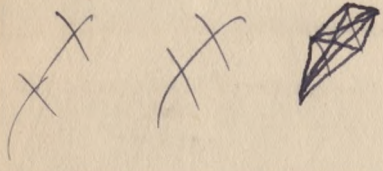
The New PHYSICS in EVERYDAY LIFE

HENDERSON

Charles' Law: volume of gas varies directly as changes 1/273 per degree centigrade.

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THE NEW PHYSICS
IN
EVERYDAY LIFE

BY

WILLIAM D. HENDERSON, PH.D.

UNIVERSITY OF MICHIGAN

LYONS AND CARNAHAN

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PREFACE

SCIENCE is ever searching for new facts and their corresponding implications and applications. Some implications at first may not seem to fit into the prevalent "picture of things," but the search for scientific truth is a fascinating task about which every high-school student should know.

In this 1935 edition of the NEW PHYSICS IN EVERYDAY LIFE there has been added at the end of the book a summary of the "Recent Investigations as to Atomic Structure and Radiations." In addition, there will be found throughout the book changes here and there in certain paragraphs. These changes are intended to provide more recent examples of known facts or to shed new light in the field of science.

New provisions for individual differences will be noted. The problems in all "Review Exercises" have been carefully arranged for three levels of ability and plainly designated by the letters *A*, *B*, and *C*. The author has intended to provide every student with a thorough and comprehensive series of review exercises which best suit his own level of development.

This edition still retains the following characteristic features which have made the NEW PHYSICS IN EVERYDAY LIFE an outstanding text: (a) *Historical Method* of approach to major subjects. (b) *Familiar Phenomena and Fundamental Principles.* Starting with familiar phenomena lying within the experience of the student, there follows a discussion of the fundamental principles involved. (c) *Abundant Illustrative and Explanatory Exercises* designed to clarify and interpret the laws and principles involved. (d) *Practical Applications.* An unusually large number of practical applications are cited and explained.

Ann Arbor, Michigan W. D. HENDERSON.

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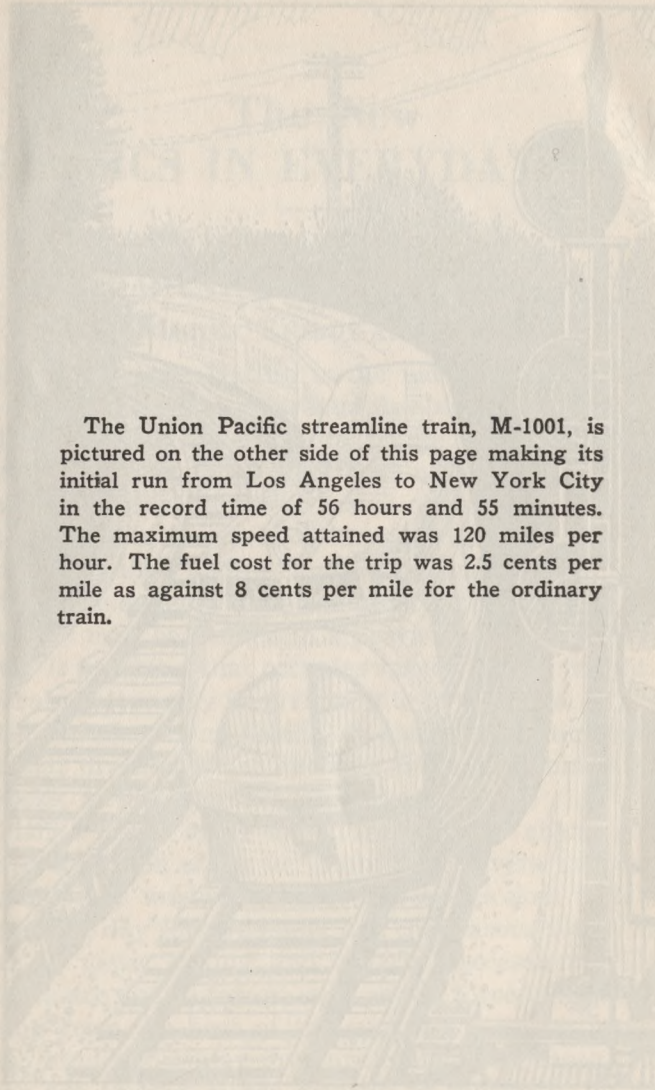
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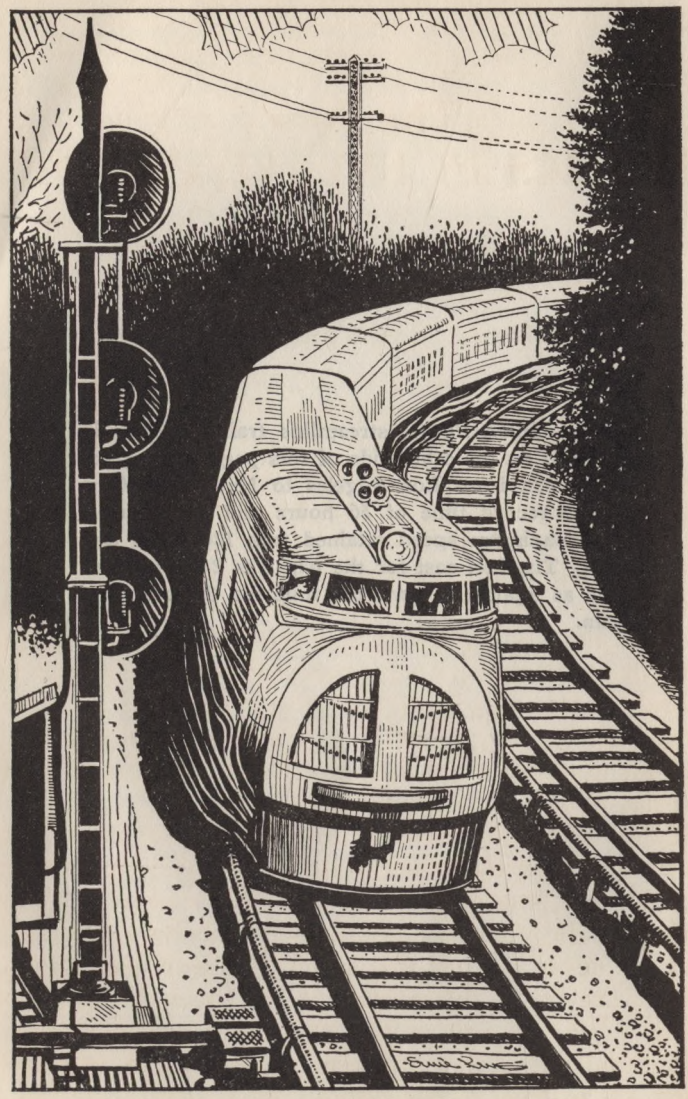
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The Union Pacific streamline train, M-1001, is pictured on the other side of this page making its initial run from Los Angeles to New York City in the record time of 56 hours and 55 minutes. The maximum speed attained was 120 miles per hour. The fuel cost for the trip was 2.5 cents per mile as against 8 cents per mile for the ordinary train.



FROM LOS ANGELES TO NEW YORK IN LESS THAN 57 HOURS!

The New
PHYSICS IN EVERYDAY LIFE

CHAPTER I

FUNDAMENTAL PHYSICAL CONCEPTS

Matter, Energy and Force

1. Introductory. Nature is the name that is given to the world which we see around us—sky and clouds, hills and valleys, fields and forests, plants and animals. Now there are two obvious facts about nature which are emphasized by even the most casual observation. The first of these is that the world of our experience is made up of something called matter, and the second obvious fact is that change appears to be the order of nature.

Matter may be defined in a provisional way by saying that it is anything that occupies space and has weight. In general, matter may be divided into two classes, (a) that which includes living things, as plants and animals, (b) and that division which includes inanimate things, such as rocks, metals, water, air, and so forth.

Change is the order of nature. We see changes constantly going on about us. Rain falls, rivers run to the sea, hills are washed down, rocks are broken into tiny fragments, day follows night, and the seasons follow each other in never ending and orderly procession. These changes are due to forces, such as the force of expansion arising from the heat of the sun, the bursting force of frost, wind and water pressure, as well as those forces which are exerted either directly or indirectly by man. Whenever a change of any kind occurs in nature, we may be sure that force has been exerted; and whenever a force

is exerted, energy is always expended. Back of force there is always energy.

Energy is the capacity which a body has for doing work. Whenever anything is capable of doing work in any form whatsoever, we say that it possesses energy. Thus the human body possesses energy because it is capable of doing work. Likewise the steam in an engine, the water in a mill pond, the coiled spring of a watch which has been wound, all possess energy. It should be noted, however, that energy does not necessarily imply the actual doing of work, but the capacity to do work. The term *work,* as used in physics, implies the overcoming of resistance through space.

Now while man has observed that the changes due to the forces of nature are constantly going on, he has also learned by experience that these changes do not "just happen," but that they take place in a perfectly definite and orderly manner in accordance with what are called the laws of nature. The study of these laws with reference to their meaning, operation, and application, forms an important part of that branch of learning called science. *Science* is commonly defined as classified knowledge. The biological sciences, such as botany, zoology, and physiology, deal with the growth and development of living things; the physical sciences, on the other hand—astronomy, physics, and chemistry—deal in a more general way with matter and energy, and the result of interacting forces.

Physics is the science which treats of matter and energy; it includes the subjects of Mechanics, Heat, Electricity, Sound, and Light.

2. Physical and Chemical Changes. Matter may be changed in two ways, physically and chemically. A *physical change* is one which does not alter the identity of the substance. The tearing of a piece of paper is a physical

change. The form of the paper is altered by the tearing, but not the substance; each piece is still paper. A *chemical change* is one which alters the identity of the substance. The burning of a piece of paper, for example, is a chemical change. In combustion (burning) the oxygen (O) of the air unites with the carbon (C) of the paper, forming a gas called carbon dioxide (CO_2). The ash which remains is composed of mineral constituents of the paper and some unconsumed carbon.

Sometimes changes occur which are both physical and chemical, as illustrated in the chewing of a piece of bread. The crushing of the bread with the teeth is a physical change; the reaction of the saliva with the starch of the bread, changing it to sugar, is a chemical change.

EXERCISE. **1.** What changes (physical, chemical, or both) are involved in (a) the breaking of a piece of glass? (b) the burning of a lump of coal? (c) the melting of a piece of ice? (d) the cooking of food? (e) the dissolving of a metal in acid?

3. The Structure of Matter. In the light of our knowledge today we think of matter as composed of small particles which are in constant motion. These particles, which are so small as to be invisible under the most powerful microscope, do not rest one upon another as do the bricks in a wall, but they are in constant motion, striking against each other, and bounding back and forth with great rapidity. In view of this explanation, the question at once arises, How does matter retain its form? If the particles of a piece of crayon, for example, are in constant motion, how does it happen that the crayon retains its shape? The ancients supposed that the particles of which matter is composed were held together by hooks or claws; we of today assume that they are held together by forces which allow considerable freedom of motion yet which retain the particles within the form of the body.

Molecules. By way of illustration, let us consider the division of any substance, such as water, into its ultimate particles. A glass of water is composed of many drops; each drop may be divided into smaller and smaller parts until, it may be conceived, we come to the smallest particle that can exist as water. This ultimate particle of water is called the molecule. A molecule is the smallest particle of a substance that can exist and retain its identity.

Atoms. If now we carry the division still further, as can be done by chemical means, we find that the molecule may be broken up into still smaller particles called atoms. The molecular symbol for water is written H_2O. This means that the molecule of water consists of three atoms —two atoms of hydrogen (H) and one of oxygen (O).

Electrons. There are many reasons which lead us to believe that atoms in turn are made up of smaller electrically charged particles called electrons, which will be discussed later in connection with our study of electricity.

Some idea of the relative number and size of atoms in a given space may be obtained from the following: "A drop of water contains several thousand million million million atoms. Each atom is about one hundred-millionth of an inch in diameter. But this is not the limit. Within the atom are the much smaller electrons pursuing orbits, like planets round the sun, in a space which relatively to their size is no less roomy than the solar system."

4. Mass. One of the most important concepts associated with matter is that of mass. In order to get at the meaning of this term, let us consider two balls of the same size, one of iron and the other of wood. Aside from the materials of which they are composed, it may be noted that they differ from each other in two very important particulars. First, the iron ball is heavier than the wooden ball. And in the second place, if each be struck a blow with a mallet, it will be observed that the iron ball

appears to resist being set in motion to a much greater degree than the wooden ball. In other words, with blows of the same intensity we can knock the wooden ball much farther than the iron ball. The iron ball contains more matter than the wooden ball and therefore it is heavier, and for the same reason it offers greater resistance to change of rest or motion. This resistance to change in its condition of rest or motion is called inertia.

Mass is the quantity of matter which a body contains; it is in a sense the measure of a body's inertia.

The measurement of mass is commercially and scientifically of great importance. The most accurate method of determining mass is by means of a beam balance, by which method the mass of a given body is compared with a known mass such as a standard gram or pound.

5. The Beam Balance. A *beam balance* is an instrument for comparing masses. It consists essentially of a beam *ABC*, Fig. 1, which is supported upon a knife edge at its middle point *B*. The arms *AB* and *BC* are of equal length. Suspended from each end of the beam are scale pans M_1 and M_2. The beam is provided with a pointer *P* which moves over a scale *S*. When the balance is without load and is at rest, the position of the pointer is called the resting point or zero. To find the mass of a given body, we place it in the left hand pan of the balance. This causes the pointer to swing to the right. We then add standard masses (weights) to the right hand pan until the pointer comes back to its zero posi-

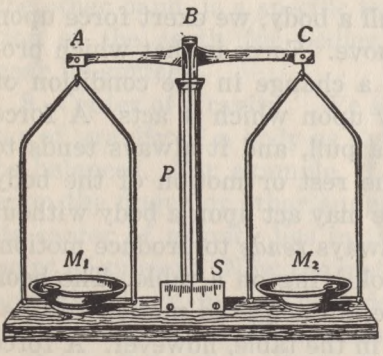


FIG. 1. Beam balance

tion. By noting the sum of the masses (weights) on the right hand pan M_2, we are able to determine the mass of the body in M_1.

The greater the delicacy of adjustment of the moving parts of the balance, the greater is the accuracy of determination.

6. Force. Among the earliest experiences of life are those which are associated with pushes and pulls upon objects, and the resulting motions of these objects. In early childhood we are undoubtedly interested primarily in the object itself, as, for example, a brightly colored ball; but a little later on this interest is transferred, in a large measure at least, to the motions which may be imparted to the ball. As experience ripens, we soon learn that in the great world about us there are innumerable pushes and pulls, and as a consequence there results a great variety of motions, other than those in which we personally take part.

Whenever we push or pull a body, we exert force upon it, and the body tends to move. *Force* is that which produces or tends to produce a change in the condition of rest or motion of the body upon which it acts. A force always implies a push or a pull, and it always tends to change the condition of the rest or motion of the body upon which it acts. A force may act upon a body without producing motion, but it always *tends* to produce motion. This is illustrated by a book lying on a table. The book exerts force upon the table, although no motion results; it tends to produce motion in the table, however. A force acting upon a body at rest may set it in motion, or, on the other hand, the application of a force may bring a moving body to rest. And again the application of a force may merely change the direction of motion of a moving body. The relation between force and motion is very well illustrated in the playing of a game of ball. The pitcher

exerts force upon the ball, producing motion; the batter exerts force upon the ball, changing the direction of its motion; the catcher exerts force in destroying its motion.

7. Gravity and Gravitation. Daily experience teaches us that all bodies tend to fall to the earth. A piece of crayon held in the fingers is at rest; if released it at once falls to the floor. Its condition of motion is changed. Now, when the motion of a body is changed, we say that a force has acted upon it. *Gravity* is the force with which the earth attracts bodies. The apple falls from the tree to the earth because of the force of gravity; rain falls from the clouds, rivers run to the sea, bodies everywhere tend to fall because of the constant pull exerted by the force of gravity.

Gravitation is the force with which every particle of matter in the universe attracts every other particle. Gravitation is a general term referring to the universal attraction which exists between all bodies; gravity, on the other hand, is a specific term referring to the attraction of the earth for bodies, usually considered at or near its surface.

8. Center of Gravity. We sometimes speak of the center of gravity of a body as the point about which it may be balanced. For example, if a meter stick be balanced upon the finger or other support about its middle point, its center of gravity lies in the stick directly above the point of support. Now the meter stick, or for that matter any other body, may be considered as being made up of a great number of mass particles, each of which is attracted to the earth by the pull of gravity; that is, the body as a whole is acted upon by a force which is equal to the sum of all the pulls exerted upon the individual particles. It is evident that when the body is balanced, as in the case of the meter stick mentioned above, the supporting force is (a) equal to the sum of all the little pulls of gravity,

and (b) it is oppositely directed to them. The *center of gravity* of a body, then, may be defined as the point of application of the resultant of all the forces of gravity acting upon the mass particles of which it is composed.

The center of gravity of a rectangular block of uniform density lies at the intersection of its diagonals; that of a circular disc, at its center of figure. In the case of bodies of irregular form the position of the center of gravity cannot thus be determined geometrically. We have to use experimental methods. In some cases the simplest way of finding the center of gravity is by the balancing method, as in the case of the meter stick. It is, however, sometimes convenient and often necessary to use the plumb line method. A *plumb line* consists of a string having a weight or "bob" attached to one end. When the bob is suspended by the string and at rest, the plumb line lies in a vertical direction; that is, it points toward the center of the earth.

9. How to Find the Center of Gravity of a Body. Suppose now that we wish to find the center of gravity of two bodies, pieces of board or metal, cut into shapes shown in Fig. 2. The method of procedure is as follows: Suspend the body in question, together with a plumb line, successively from the points *A* and *B*, noting the position of the plumb line in each case when the system is at rest. The center of gravity lies at *C*, the intersection of the lines *AA'* and *BB'*, and at a point midway between the faces. Usually the center of gravity lies

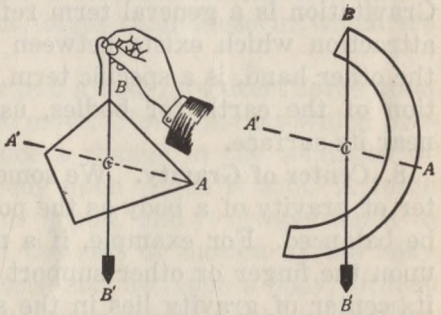


FIG. 2. Center of gravity

within the material of which the body is composed. This is not always true, however, as is illustrated by the case of the curved body, Fig. 2. Here the center of gravity C is located at a point entirely outside of the body itself. The center of gravity of a ring of uniform shape and density is at the center of the circle formed by the ring; the center of gravity of a spherical shell is at the center of the figure.

EXERCISE. **2.** With reference to the material of which the body is composed, at or near what point is the center of gravity of (a) a metal cylinder? (b) a teacup? (c) a ball bat? (d) a boat?

10. Weight. The weight of a body is the force with which it is attracted to the earth. *Weight is a force,* directed from the center of gravity of the body considered to the center of the earth.

The weight of a body depends upon two factors, (a) the quantity of matter which it contains (its mass) and (b) its position with respect to the earth. Thus, for example, for a given place on the earth's surface a whole brick weighs more than a half brick, because the first has more mass than the second. Also, a brick at the surface of the earth weighs more than if taken some distance above the surface, because the force of gravity is greater at the surface than above or below. The relation of the weight of a body to its position with reference to the earth's surface may be stated briefly as follows:

Weight at the surface. The nearer a body is to the center of the earth, so long as it remains upon the surface, the greater is its weight. Thus, a given mass will weigh more at the base of a mountain than at the top. Also, a body at or near the poles of the earth will weigh more than at or near the equator. This is due to the fact that the equatorial radius is some 13 miles greater than the polar radius, and also due in part to the increased

rotational effect upon bodies at the equator. At the equator the weight of a given mass is 5 parts in 1000 less than at the poles; that is, a mass having a weight of 1000 grams at the poles would weigh only 995 grams at the equator.



FIG. 3
Spring balance

Weight above or below the surface. The weight of a body above or below the surface of the earth is less than at the surface. A mass weighing 100 pounds at the surface of the earth will weigh less than 100 pounds if taken up in a balloon or down into a mine.

The weight of a body may be considered to be concentrated at its center of gravity.

11. The Spring Balance. The weight of a body may be determined by means of a spring balance. In Fig. 3 there is shown a laboratory balance which has a double scale. On the left it is graduated to give readings in grams and on the right in ounces. The spring balance contains a spiral spring, the upper end fastened to the frame of the instrument, and to the lower end there is attached a hook. In front of the spring there is placed a scale plate, over which the pointer moves as the spring is pulled downward.

To graduate a spring balance, that is, to fix its scale numbers, known masses (weights) are successively attached to the hook. For example, a 50-gram mass is attached to the hook, and the pointer moves down to a given position; this is marked 50 on the scale plate. Then a 100-gram mass is suspended from the hook and the resting point is marked 100, and so on. The ounce scale, which appears on the right, may be obtained in a similar

manner by attaching successively to the hook masses of 1 ounce, 2 ounces, 3 ounces, and so on. It is also possible to compute the ounce scale from the left hand scale, 100 grams being equivalent to 3.527 ounces. The method of using successive weights of increasingly greater masses, as explained above, is employed when a high degree of accuracy is desired. Ordinary commercial balances are usually graduated by the use of a single weight.

Now since the pull of gravity may vary slightly from place to place, it follows that spring balances which are graduated at places having different altitudes (at Washington, D. C., and Denver, Colorado, for example) will not have exactly the same scale displacements. To obviate this discrepancy, balances are graduated in terms of the force of gravity at some fixed level on the earth's surface. The place selected for such graduation is at or near the sea level, latitude 45°.

A spring balance may not only be used for measuring the weight of a body, that is, the force of gravity acting upon it, but also for measuring the magnitude of a force applied in any direction whatsoever. For example, if we pull on the hook of a spring balance until the pointer comes to the 200-gram mark, we say that we are exerting a force of 200 grams.

The spring balance is sometimes called a *dynamometer*, that is, a force measurer. By the proper selection of the size and material of the spring, spring balances may be constructed suitable for the measurement of forces ranging in magnitude from grams and ounces to pounds and even tons.

12. Different Uses of the Term "Weight." In the study of physics the term *weight* occurs many times, and it is sometimes used in three different senses. In order to avoid unnecessary confusion in the mind of the student, it is important that these different uses of the word be

explained. (a) First, the term "weight" is used to designate the pull of gravity upon a body, as explained in Art. 10. When used in this sense weight is a force.

(b) Second, the word is used to designate an object, as a piece of metal. For example, we may say, "Put the weight on the scale pan of the balance." Every laboratory is provided with a set of "weights."

(c) In the third place, we sometimes use the word *weight* when we mean *mass*. For example, when we weigh a body by means of a beam balance, we frequently say, "Its weight is 10 grams." What we really mean here is that its mass is 10 grams, because a beam balance is used to compare, that is, to determine masses.

13. Distinction between Mass and Weight. It is very important that a distinction be made between the mass of a body and its weight. Mass refers to the *quantity of matter* in a body; weight, to the *force* with which the earth attracts a body.

If an object is moved from one place to another, its mass is not affected thereby; its weight, on the other hand, may be changed, since the weight of a body is determined by the force of gravity acting upon it; and the force of gravity for a given mass differs slightly for different places upon the earth's surface. For example, if a body having a given mass be weighed by means of a sensitive spring balance at New York City (sea level) and again on the same balance at Denver, it will be found to weigh less at the latter place than at the former, because the force of gravity is considerably less at Denver than at New York. The mass of the body in question is exactly the same in New York as at Denver, although its weight is different.

A beam balance, to repeat, is used to determine masses, and such determinations are independent of the location of the place where the experiment is performed. A

spring balance, on the other hand, is primarily an instrument for measuring force. The force may be that of gravity (weight), or it may be a force in any direction we may choose. In certain commercial transactions and household measurements spring balances are now largely used for weighing, because of the ease with which the operation may be performed. It should be remembered, however, that these instruments really measure forces, which are assumed to be practically proportional to the masses which are placed upon their scale pans. If a grocer's spring balance, used for weighing sugar, coffee, etc., is calibrated at sea level and then taken to a place above sea level, it is evident that somewhat more than a pound will have to be added to the pan to bring the pointer to the one-pound mark on the scales. The difference between the masses considered in the two cases mentioned above is, however, so slight that in ordinary commercial transactions it is usually negligible.

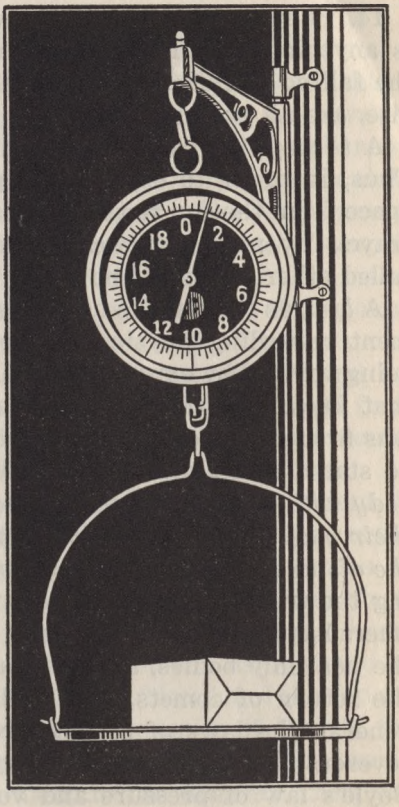


FIG. 4.—Household balance

A household balance used for weighing commodities put up in small packages is shown in Fig. 4.

14. Phenomenon, Theory, Law. A *natural phenomenon* is anything occurring in nature. The rising of smoke, the falling of rain, the sound of a whistle, the flying of a kite, are all examples of natural phenomena.

A *theory* is a reason put forth to explain phenomena. Thus, in order to explain the passage of light through space from the sun to the earth, we assume that the light travels by means of waves through an elastic medium called ether. This is known as the wave theory of light.

A *law,* as the term is used in science, is a definite statement regarding physical phenomena, and is capable of being verified by experiment. A good example of a law is that known as the universal law of gravitation, which was first announced by Sir Isaac Newton. This law may be stated as follows: *Every body attracts every other body with a force directly proportional to the product of their masses and inversely proportional to the square of the distance between their centers of gravity.* By assuming the truth of this general law of gravitation, astronomers have been able to describe accurately the motions of the heavenly bodies, as, for example, to predict eclipses, the return of comets, and to discover new planets. In general, the laws of science are named after their discoverers, as for example, Newton's law of gravitation, Boyle's law of pressure and volume, and Ohm's law of current strength.

An *experiment* is a question put to nature, the answer to which is determined by accurate tests and measurements. All experimentation is based upon the assumption of the constancy of nature; that is, we assume that under the same conditions nature always acts in the same manner. In order to carry out experiments, it is necessary to have fixed and definite standards of measurement. We

shall, therefore, now take up for consideration our fundamental units of measurement.

Fundamental Units of Measurement

15. Introductory. Since the study of physics includes not only the observation and classification of physical phenomena, but also the measurement of these phenomena, it is essential that we have at the very outset a clear understanding of the units employed. The *fundamental units* for all physical measurements are those of *length, mass,* and *time.* All the units which are used in the measurement of physical quantities, such as the foot pound (unit of work), the horse power (unit of power), the ampere (unit of current strength in electricity), and the candle power (unit of illumination) are determined from and may be expressed in terms of these fundamental units named above, namely, length, mass, and time.

Early in the history of the race the measurement of length, volume, mass, and time became necessary in order that people might carry on trade and other forms of commercial intercourse. The first units employed were, naturally, those derived from parts of the human body, and such commonly observed phenomena as occurred at regular intervals. For example, the "foot" and the "span" were quite naturally selected as the units of length, the former being a length equal to the average adult human foot, and the latter equivalent to the distance from the end of the thumb to the tip of the longest finger of the outstretched hand.

Early concepts of mass were expressed in terms of weight as measured by the muscular effort required to lift a given body. Even to the present day people frequently estimate the mass of a body by "hefting" it, that is, by lifting it and noting the amount of muscular effort put forth.

The passage of time then, as now, was recorded approximately by the seasons, the phases of the moon, and by the recurrence of day and night.

16. Household Measurements. Before the days of "Household Economics" and "Domestic Science" practically all measurements involving household processes were made by the rule of thumb method. Even to this day we are familiar with such vague and unscientific terms as a "handful of flour," a "pinch of salt," and "a few drops of oil." The use of the terms "spoonful" and "cupful" indicate a degree of improvement in the line of accuracy, but the day is at hand when, at the behest of science and economy, these common units of measurement will have to give place to more accurate and scientific terms. In factories, hospitals, and large hotels where food is prepared on a large scale, masses and volumes are measured with precision, temperatures determined accurately, and the time factor in each process is carefully controlled. And in these days when science is being extended to every phase of common life, there is no valid reason why similar methods, which make for efficiency and economy, should not be applied to most, if not all, of the ordinary household processes. Suitable apparatus for the measurement of mass, volume, and temperature can now be procured at a small cost, and no household in which economy and health are of prime importance should be without these appliances.

17. Industrial Processes. Modern industry owes its efficiency in no small measure to the fact that careful measurements today control processes, fix dimensions, determine proportions, and standardize products. The accurate measurement of the parts of machines or accessories makes it possible to manufacture these parts separately and to assemble them without fitting. In many industries, also, efficient operation is dependent upon measured con-

trol of heat, electric current, or power. Steel should be "quenched" at a certain stage in cooling, bread should rise at a given temperature, and similar account must be made of many other processes in order to secure the best results. Properties of materials such as iron and steel, paints and oils, cement and building tile, and many other substances may be accurately determined, thus giving a means of controlling quality and establishing fixed standards. Indeed, we may well say that in the modern world measurement guides the processes of all industries and keeps their output up to fixed standards.

18. Fundamental Legal Standards. The fundamental units of length, volume, capacity, and mass of the United States are those of the metric system, which was legalized by Act of Congress in 1866. While it is true that in this country we commonly employ the inch, the foot, the yard, and ounces and pounds as units of length and mass, it should be born in mind that in the United States these units are legally defined not in terms of the corresponding British standards but in terms of the metric system.

The metric system was first introduced by the French about the year 1793, and has since been adopted, in whole or in part, by most of the civilized countries of the world. The meter was originally intended to be equivalent to one ten-millionth of an earth quadrant, that is, one-fourth of a great circle of the earth; and the kilogram was intended to be equivalent to the mass of one liter of pure water. As it was found impossible, however, to determine exactly these quantities, there were arbitrarily chosen as standards a meter and a kilogram which are approximately equal to the theoretical values determined upon by the originators of the system.

The meter and kilogram which were finally chosen as standards were made of platinum and are kept in the

Palace of the Archives at Paris. They are known as the "Standards of the Archives."

19. The International Metric Standards. In 1872 an International Conference of Weights and Measures was called at Paris. The object of this conference was to consider the question of International Standards. Thirty countries responded, the United States being among the



FIG. 5.—International Laboratory at Sèvres, near Paris, France, where the International Metric Standards are kept

number. At this meeting three things were accomplished: (a) an International Bureau of Weights and Measures was organized; (b) an International Laboratory, Fig. 5, located at Sèvres, near Paris, was established, and (c) the construction of a number of prototype standards, similar to the Standards of the Archives, was authorized. As a result of this conference a number of standard meters and kilograms were constructed of an alloy of 90 per cent platinum and 10 per cent iridium, each instrument being as nearly as possible an exact duplicate of the corresponding Standards of the Archives. That meter and that

kilogram which most closely corresponded to the meter and the kilogram of the Archives were chosen as the International Standards, and were deposited in the International Laboratory, where they are now kept for reference. The remaining standards were disposed of by lot to the various countries represented, the United States drawing two meters, Nos. 21 and 27, and two kilograms, Nos. 4 and 20. As has already been stated, these standards were brought to this country and are now kept in the Bureau of Standards at Washington, Fig. 6.

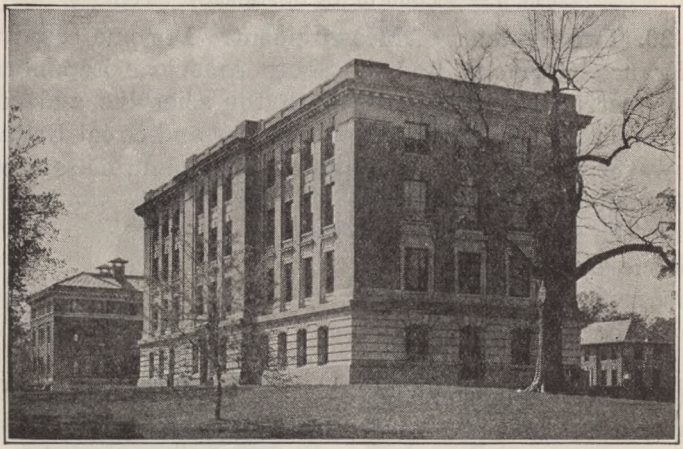


FIG. 6.—Bureau of Standards building, Washington, D. C., in which the U. S. metric standards of length and mass are kept

A very good idea of the appearance of the U. S. standard meter and kilogram may be obtained from Figs. 7, 8, and 9, which are copies of actual photographs. Fig. 7 shows a longitudinal section of U. S. prototype meter No. 27, and Fig. 8 a cross section of this standard meter, actual size. Fig. 9 is reproduced from a photograph of standard kilogram No. 20. The standard kilogram is

cylindrical in shape, rests upon a circular supporting base, and is enclosed within two glass bell jars, as shown.

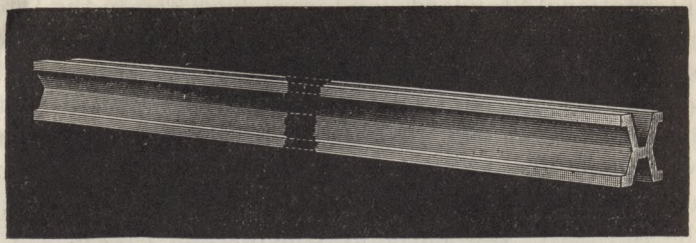


FIG. 7.—Section of U. S. standard meter

20. Two Systems of Measurement. In some countries (France, Germany, and the South American Republics), the metric system is used both for scientific and commercial purposes. In the United States and Great Britain, however, two systems of units of length and mass are in common use, the metric and the English. The metric unit of length is the meter; the English unit, the yard. The metric unit of mass is the kilogram; the English unit, the pound.

It is important for the student to note that a distinction must be made between the English units of the United States and those of Great

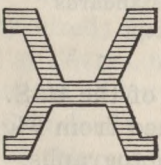


FIG. 8.—Cross section of U. S. standard meter

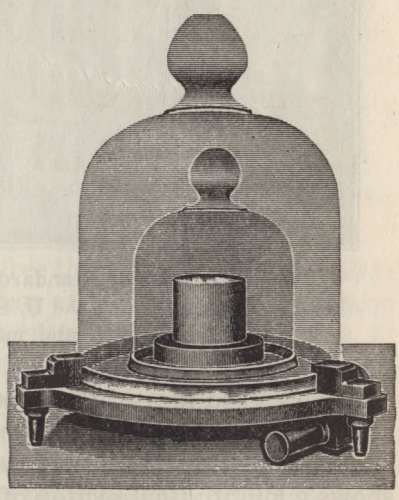


FIG. 9.—U. S. standard kilogram

Britain. Our English units, the inch, the foot, the yard the pound, and so forth, came historically from those of England, but in actual practice today they are derived from our national metric standards at Washington. For example, the U. S. inch as defined by Act of Congress is as follows:

$$1 \text{ inch (U. S.)} = \frac{1}{39.37} \text{ of a meter.}$$

In Great Britain, on the other hand, the inch is defined as 1/36 of a standard British yard. The standard British yard and pound, together with the metric standards of that country, are kept at the office of the Exchequer in London.

The U. S. inch is slightly longer than the British inch, as shown by the following: 1 U. S. inch = 1/39.37 m = 2.54 cm; 1 British inch = 2.538 cm.

Since the English units of this country are defined by law in terms of the metric system, we shall consider first the metric units of length, volume, capacity, and mass, after which we shall define the corresponding English (U. S.) equivalents.

21. Metric Units of Length. The metric unit of *length* is the meter. A *meter* is the distance between two cross marks on a platinum-iridium bar which is kept in the Bureau of Standards at Washington, the measurement being made at 0° C. In designating fractions of the meter, we use the Latin prefixes, deci, centi, milli; multiples, the Greek prefixes, deka, hecto, kilo. The divisions and multiples of the meter are:

Fractions		*Multiples*	
1 decimeter (dm)	= .1 meter	1 dekameter (dkm)	= 10 meters
1 centimeter (cm)	= .01 meter	1 hectometer (hm)	= 100 meters
1 millimeter (mm)	= .001 meter	1 kilometer (km)	= 1000 meters

The three most commonly used units of the above list are the centimeter, the millimeter, and the kilometer.

The relation between the inch and the centimeter is shown graphically in Fig. 10. For additional data on Units of Length, Area, and Volume, see Tables I and II, Appendix.

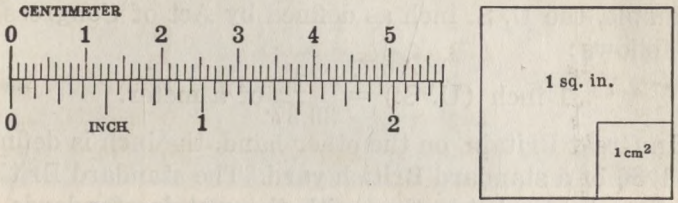


FIG. 10.—Comparison of centimeter and inch

22. Metric Units of Volume and Capacity. The metric unit of volume is the cubic meter (m^3). The fractional units of volume are the cubic decimeter (dm^3) and the cubic centimeter (cm^3).

The metric unit of capacity is the liter. A *liter* is 1000 cubic centimeters. The Bureau of Standards defines the liter as the volume of one kilogram of air-free distilled water at 4° Centigrade. The reason for thus defining the liter in terms of the volume of a kilogram of air-free distilled water at 4° C, instead of defining it directly as 1000 cubic centimeters, is because of the convenience of calibrating glass flasks and similar vessels. Suppose that we wish to calibrate a flask similar to the one shown in Fig. 11. Place the empty flask on the left-hand pan of a suitable beam balance and add weights to the right hand pan until the pointer comes back to zero. Then place upon the right hand pan a standard mass of one kilogram. Now carefully pour into the flask distilled water of the required temperature until the balance is again in equilibrium. The flask now contains one kilogram of water. Make a circular mark

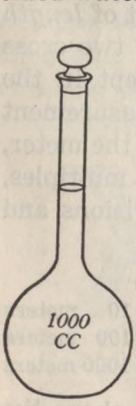


FIG. 11
Liter flask

on the neck of the flask, indicating the surface of the water. The flask is now calibrated in terms of cubic centimeters. That is, its capacity when filled with a liquid up to the circular mark is 1 liter or 1000 cc. A 500 cc "graduate" is shown in Fig. 12.

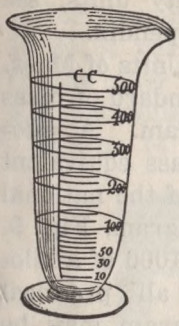


FIG. 12
Graduate

One one-thousandth of a liter is a milliliter (ml). The milliliter is usually designated by the symbols cc, usually read "see-see."

1 liter = 1000 cc
1 cc = 1/1000 liter

In general we use the term "volume" to refer to the size of a body. A rectangular block having a length of 20 cm, a width of 10 cm, and a height of 5 cm has a volume of 1000 cm^3, that is, $20 \times 10 \times 5 = 1000$ cm^3. On the other hand, when we use the term "capacity" we think of how much a vessel will hold. For example, the capacity of the flask shown in Fig. 11, when filled up to the mark, is 1 liter or 1000 cc. Volume in cubic centimeters is in general designated by the symbols cm^3; capacity in cubic centimeters, by the symbols cc. This is a mere convention, which holds, however, in most physical and chemical laboratories.

The U. S. English equivalents are:

1 liter = 0.908 quart (dry measure)
= 1.0567 quarts (liquid measure)
1 dry quart = 1000/0.908 = 1101.3 cc
1 liquid quart = 1000/1.0567 = 946.3 cc

It thus appears that a dry quart has a greater capacity than a liquid quart. The comparative capacities of these units are shown graphically in Fig. 13. Certain dry commodities, such as beans, peas, etc., which are supposed to be sold by the quart dry measure are

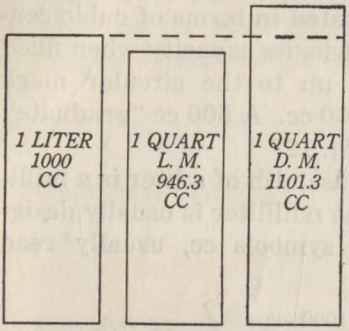


FIG. 13.—Comparative volumes of liter, liquid quart, and dry quart

sometimes fraudulently sold liquid measure. For additional Capacity units, see Table III, Appendix.

23. Metric Units of Mass. The U. S. standard of mass is the kilogram. A *kilogram* is a mass equivalent to the mass of the national standard kilogram, Fig. 9. A gram is 1/1000 of a kilogram. For all practical purposes a gram may be considered to be equivalent to the mass of a cubic centimeter of air-free distilled water. The divisions of the kilogram and gram are given in the following table:

1 gram (g)	= 1/1000	kilogram (kg)
1 decigram (dg)	= 1/10	gram
1 centigram (cg)	= 1/100	gram
1 milligram (mg)	= 1/1000	gram
1 kilogram	= 2.2046	pounds
1 gram	= 15.432	grains

For additional mass units, see Table IV, Appendix.

According to the rules of the U. S. Mint, the following pieces of money, as legalized by Act of Congress, are coined in terms of the metric units of mass. The mass of a

5 cent piece (nickel)	=	5.0 grams
10 cent piece (dime)	=	2.5 grams
25 cent piece (quarter)	=	6.25 grams
50 cent piece (half dollar)	=	12.5 grams

The fineness of gold and the weight of precious stones are measured in terms of the "karat" and the "carat" respectively.

1 karat (fineness of gold) = 1/24 by weight of pure gold
1 carat (in precious stones) = 200 milligrams

For example, a ring of 18 karats fineness contains 18/24 of pure gold; 24 karats fine signifies pure gold. On the other hand, a 5-carat diamond is one that weighs 5×200 milligrams = 1 gram.

In Fig. 14 there is shown an analytical beam balance

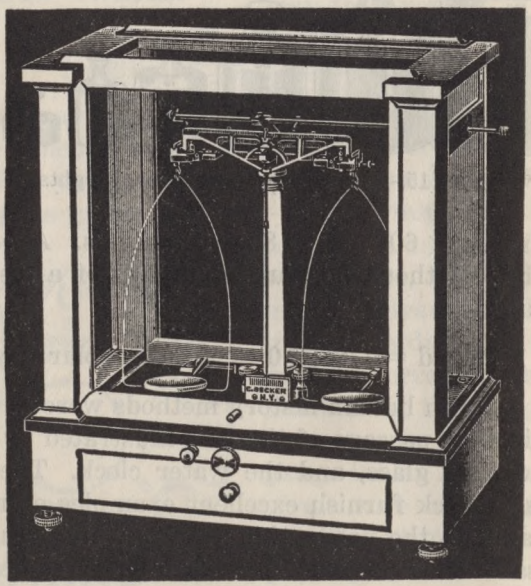


FIG. 14.—Analytical balance

used for making accurate determinations of mass, and in Fig. 15 a set of standard gram masses (weights) ranging from 1 to 100 grams.

24. The Unit of Time. The interval between two successive passages of the sun across a given meridian is a solar day. Solar days vary in length throughout the year, being longest in summer and shortest in winter. A *mean*

solar day is the average of all the solar days taken throughout the year. The time recorded by watches and clocks is mean solar time.

The unit of time is the second. In a mean solar day



FIG. 15.—Set of standard gram weights

there are $24 \times 60 \times 60 = 86{,}400$ seconds. A *second* is one eighty-six-thousand-four-hundredth of a mean solar day; that is

$$1 \text{ second} = 1/86{,}400 \text{ of a mean solar day}$$

Very early in human history methods were devised for measuring the passage of time, as illustrated by the sundial, the hour glass, and the water clock. The modern watch and clock furnish excellent examples of manufacturing skill in the production of relatively accurate and at the same time very cheap recording instruments.

For railroad engineers, signal officers, and other public officials it is extremely important that timepieces be carefully adjusted and frequently tested. The chronometer is a specially jeweled and very carefully adjusted clock, usually put up in box form, which is used as a standard of reference. In modern practice time signals are sent out once a day from the National Observatory at Washington by radio.

25. Angular Measure. The units of angular measure

are the *degree* and the *radian*. In all ordinary measurements of angles, the degree is the unit used; in certain mathematical computations, the radian is used. A degree is 1/360 of a circle. In a circle there are 360 degrees (360°) or 2π radians. We may then write

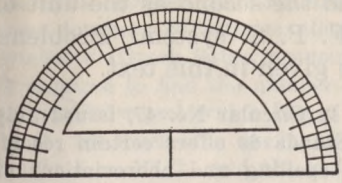


FIG. 16.—Protractor

$$1 \text{ circle} = 360 \text{ degrees} = 2\pi \text{ radians}$$

$$1 \text{ radian} = \frac{360}{2\pi} = \frac{360}{2 \times 3.1416} = 57.3°$$

Angles may be measured by means of a *protractor*, Fig. 16, which is a semi-circular scale divided into 180°.

26. Wire Gauge. The size of a wire is usually designated by its gauge number. To determine the diameter of a given sample of wire, a wire gauge, Fig. 17, may be used to advantage. In order to find the diameter in millimeters corresponding to a number of the gauge, it is necessary to refer to a wire gauge table. For example, a given wire which just fits hole No. 14 in the gauge has a diameter of 1.628 mm, as shown by Table V, Appendix, in which diameters are given in millimeters and in mils. A *mil* is 1/1000 of an inch.



FIG. 17.—Wire gauge

27. C. G. S. and F. P. S. Systems. When in making a physical measurement we use the centimeter as the unit of length, the gram as the unit of mass, and the second as the unit of time, we speak of the system of units thus employed as the C. G. S. system. Likewise, when in the

English System we use the foot as the unit of length, the pound as the unit of mass, and the second as the unit of time, we speak of it as the F. P. S. system. Problems involving both systems will be given in this text.

28. Spelling and Abbreviations. In Circular No. 47, issued July 1, 1914, the National Bureau of Standards offers certain recommendations with reference to the spelling and abbreviations of units. Among the rules adopted by the Bureau are the following:

1. "The period is omitted after abbreviations of metric units, while it is used after those of the customary system." For example, we should write *m* for meter, *cm* for centimeter, *mm* for millimeter, and so on; in the case of units of the English system, however, we write *ft.* for foot, *lb.* for pound, *cu. ft.* for cubic foot, and so on. This notation is followed throughout this text. Some authors, however, cling to the old practice of using the period after both metric and English abbreviations.

2. "The exponents ^{2} and ^{3} are used to signify area and volume respectively in the case of the metric units, instead of the longer prefixes *sq. cm.*, or *cu. cm.*" Thus, for example, we write 10 cm^2 and 70 cm^3 instead of 10 sq. cm. and 70 cu. cm. The special abbreviation *cc* for both cubic centimeter and milliliter is in general use.

3. "The use of the same abbreviation for both singular and plural is recommended. Thus, *cm* may stand for centimeter or centimeters." For example, we write 45 cm, never 45 cms.

29. Some Important Equivalents and Equations. In the solution of problems which are given in this text, the following equivalents and equations may be of value. The numbers given in parentheses are the approximate values which should be used in all ordinary computations.

$\pi = 3.1416$; $\pi^2 = 9.8696 = (9.87)$; $\sqrt{\pi} = 1.7724$
Circumference of a circle $= 2\pi r$
Area of a circle $= \pi r^2$
Lateral area of a cylinder $= 2\pi rh$
Volume of a cylinder $= \pi r^2 h$
Area of a sphere $= 4\pi r^2$
Volume of a sphere $= 4/3\pi r^3$
Mass of 1 cc of distilled water at 4° C = 1 gram
Mass of 1 cu. ft. of water = 62.4 lb.
Mass of 1 gallon of water = 8.3 lb.

In the solution of problems involving the factor 3.1416, which is the ratio between the circumference of a circle and its diameter, some teachers of physics allow their students to substitute the symbol π (Greek letter, pronounced *pi*). For example, suppose it is required to find the number of pounds of water contained in a cylindrical vessel which has a diameter of 4 ft. and a height of 10 ft., the mass of a cu. ft. of water being taken as 62.4 lb. The height $h = 10$ ft., and the radius $r = 2$ ft. The volume then $= \pi r^2 h = \pi \times 4 \times 10 = 40\pi$ cu. ft. Hence the mass of the water $= 40\pi \times 62.4 = 2496\pi$ lb. If desired, this value may be expressed as $2496 \times 3.1416 = 7841.4$ lb.

For additional data on Equivalents, see Table VI, Appendix.

SUMMARY

1. Matter and Energy. Matter is anything that occupies space and has weight. Mass is the quantity of matter which a body contains. Energy is the capacity which a body has for doing work.

2. Physics is the science which treats of the related phenomena of matter and energy.

3. Energy and Force. Whenever energy is expended, force is exerted. Force is of the nature of a push or a pull. Force is that which tends to change the condition of rest or motion of a body.

4. Gravity and Gravitation. Gravity is the force with which the earth attracts bodies toward its center. Gravitation is the force with which every body in the universe attracts every other body in the universe. Gravity is a special term applying to the earth; gravitation is a general term applying to the universe.

5. Weight is the measure of gravity; it is a force directed from the body to the center of the earth.

6. Phenomenon, Theory, Law. A phenomenon is anything occurring in nature. A theory is a reason put forth to explain phenomena. A law is a definite statement regarding phenomena with respect to cause and effect. Physical laws may be verified by experiment, and may be expressed in terms of mathematic equations.

7. The Law of Gravitation. Any two bodies in the universe attract each other with a force which is directly proportional to the product of their masses and which is inversely proportional to the square of the distance between their centers of gravity.

8. Fundamental Units of Measurement. The three fundamental units for all physical measurements are those of length, mass,

and time. In the United States the legal fundamental units of length and mass are those of the metric system; the English units (foot and pound) are defined in terms of the corresponding metric units.

9. The meter is the legal unit of length. A meter = 39.37 U. S. inches. That is, 1 U. S. inch = 1/39.37 meter. 1 m = 10 dm = 100 cm = 1000 mm.

10. The metric unit of volume is the cubic meter. The fractional units of volume are the cubic decimeter, and the cubic centimeter. 1 liter = 1000 cubic centimeters. 1 cc = 1/1000 liter.

11. The metric unit of mass is the kilogram. 1 g = 1/1000 kg.

12. The unit of time is the second. 1 second = 1/86,400 of a mean solar day.

13. Metric and English Equivalents. For purposes of making comparisons between metric and English units, the following actual and approximate equivalents should be kept in mind:

1 meter = 39.37 inches = approximately 1 yard

1 inch = 2.54 cm.

1 liter = 0.908 quart (dry measure) = 1.0567 quarts (liquid measure) = approximately 1 quart

1 kilogram = 2.2046 pounds = approximately 2.2 pounds

REVIEW EXERCISES

A

1. Wherein do the physical sciences differ from the biological? What subjects are included in the subject of physics?

2. What is a physical change? Give some examples of physical changes other than those mentioned in Art. 2.

3. What is a chemical change? Give some examples.

4. Give an example of a change which is both physical and chemical in its nature, other than those mentioned in the text.

5. Explain the distinction between the molecule and the atom, using for illustration the water molecule.

6. What is mass, and how may it be measured?

7. Define force, and give an example (a) of a force which acts upon a body without producing a noticeable change in the motion of the body; (b) a force which changes the condition of rest or motion of a body; (c) a force which changes the direction of the motion of the body upon which it acts.

8. Explain the distinction between gravity and gravitation.

9. Define center of gravity, and tell where it is located (a) in

a rectangular block of uniform density, (b) in a circular disk, (c) in a sphere.

10. What is weight, and what are the factors upon which the weight of a body depends?

11. If a body is carried from sea level to the top of a mountain, (a) how is its weight affected? (b) its mass?

12. Explain why a given body at the poles of the earth weighs more than at the equator.

13. Explain the meaning of the words "weight" as used in the following sentence: "By means of the spring balance find the weight of the weight on the table."

14. Explain fully the distinction between mass and weight.

15. (a) What do we measure when we use a beam balance, the mass of a body or its weight? (b) What do we measure when we use a spring balance?

16. The U. S. units of length, volume, and mass are defined in terms of what standards? Where are those standards kept?

17. A meter is equivalent to how many decimeters? centimeters? millimeters? What is a kilometer?

18. A meter is equivalent to how many U. S. inches? Define the inch in terms of the meter.

19. Measure the length of this page and express your result in (a) centimeters, (b) inches.

20. A liter is equal to how many cubic centimeters? In chemical laboratories the term "milliliter" is sometimes used instead of cubic centimeter. So far as volume is concerned, is there any difference between a milliliter and cubic centimeter?

21. The inside dimensions of a cylindrical vessel are as follows: Height 20 cm, diameter 10 cm. What is the capacity of the vessel in cc?

22. What is a kilogram? Where is the U. S. standard kilogram kept?

23. A rectangular vessel has the following inside dimensions: Length 25 cm, width 10 cm, height 8 cm. This vessel is filled with distilled water at 0° C. What is the mass of the water in grams? kilograms?

B

24. A straight line drawn from the vertex of a triangle to the middle of the opposite side is called a median line. It may be

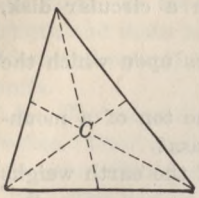


FIG. 17*a*
Center of gravity at intersection of median lines

shown mathematically that the center of gravity of a triangle lies at the point of intersection of its median lines, Fig. 17*a*. How may this be proved experimentally in the case of a triangular piece of board of uniform thickness and density?

25. Under what conditions and assumptions may a spring balance, as for example a grocer's balance, be used for estimating mass?

26. During the war the American soldiers in France (where only metric units are used) were instructed to consider a kilometer as equivalent to ⅝ of a mile. In translating 20 kilometers into miles by the use of this ratio, what actual error was made?

27. A rod is 16.5 feet. In certain states the legal width of a country highway is 4 rods. This is equivalent to how many feet? meters?

28. At the 1912 Olympiad held at Stockholm, Sweden, the 100 meter dash was made in 10.6 sec. At this rate, find the average speed of the runner in (a) feet per second, (b) miles per hour.

29. The great cannon used by the Germans in the reduction of the Belgian forts were known as "42 centimeter guns." That is, the diameter of the bore of these guns was 42 cm. Find the diameter of the bore in inches.

30. A metal cylinder has a mass of 2.5 kilograms. What is its mass in pounds?

31. In an avoirdupois pound there are 7000 grains. This is equivalent to how many grams?

32. What is the mass of $10 in nickels? dimes? quarters?

C

33. The great Kohinoor diamond, which is in the possession of the British Crown, has a mass of 125 carats. What is its mass in grams?

34. A given cylinder has a height of 20 cm, and a diameter of 10 cm. Find the area of its surface, including both ends, in cm^2.

35. The diameter of a sphere is 1 ft. 8 in. Find (a) its area in square inches, (b) its volume in cu. ft.

36. Find the ratio of the volume of a cylinder having a height of 10 cm and a diameter of 10 cm to the volume of the greatest sphere that can be cut from this cylinder.

NOTE: For additional problems, see Appendix.

start

CHAPTER II

MECHANICS OF FLUIDS

Pressure in Liquids

30. Introductory. In order to understand what follows in this and succeeding chapters, it will be necessary at the very outset for the student to get clearly in mind the meaning of certain terms, such as Mechanics, Fluids, Density, and Pressure.

We are now about to take up the study of Mechanics of Fluids. The term Mechanics, as originally used, referred primarily to the study of machines. It has, however, come to have a much wider meaning, referring in a general way to the effect of forces on bodies. The force exerted by a locomotive in pulling a train of cars, that exerted by the typist in operating a typewriter, that exerted by water in a tank on the bottom and sides of the containing vessel, and that exerted by the downward push of the atmosphere upon bodies on the surface of the earth are all illustrations of different phases of mechanics. *Mechanics* then may be defined as that subject which deals with the operation of forces upon bodies and the effects produced by these forces. For example, a study of the mechanics of the automobile would include not only a study of the essential moving parts, such as levers, shafts, cranks, gears, and so on, but also a study of the effects produced on these parts by the application of forces exerted by the hand (in steering), the foot (in applying the brakes), and the expanding gas in the cylinders.

Mechanics of fluids deals primarily with the effects of forces which are exerted on or by liquids and gases. *Hydrostatics* treats of the pressure of liquids at rest. *Hydraulics* treats of liquids in motion. *Pneumatics* treats of the pressure and motion of gases.

31. Solids and Fluids. We sometimes think of matter as occurring in three states—solid, liquid, and gaseous, as illustrated by earth, water, and air. A more specific classification, however, would assign matter to two general conditions, namely, solid and fluid. The fundamental property of a fluid is that it tends to flow. In every day experience we have, in general, no difficulty in distinquishing between solids and fluids. We do not hesitate to classify the pen with which we write as a solid and ink as a fluid. Nevertheless, in a great many cases no sharp dividing line can be drawn between solids and fluids. Indeed there is a continuous graduation from the most rigid solids to the rarest gases. Steel is a rigid solid; syrup is a viscous fluid; gasoline is a highly mobile fluid, which upon evaporation becomes a highly explosive gas when mixed with air. Some substances, such as pitch, have the properties of both a solid and a fluid. If struck a sharp blow with a hammer, a lump of pitch will break like glass; if left to itself, however, it will flatten out of its own weight and will tend to flow.

A *fluid* may be either a liquid or a gas. A *liquid* conforms to the shape of the containing vessel and always has a definite surface. A *gas,* on the other hand, has no definite surface and it always tends to fill completely the containing vessel. A *vapor* is a portion of matter in a gaseous state. When, for example, we speak of water vapor in the air, we refer not to mists or fogs but water in a gaseous condition. There is always a certain amount of invisible water vapor in the air even on the clearest day. Mists, fogs, and clouds are not vapors; they con-

sist of collections of tiny particles of water suspended in the air.

32. **Density.** *Density* is mass per unit volume. Everybody knows that a brick is heavier than a piece of wood of the same size. The density of brick (that is, its mass per unit volume) is greater than that of wood. Consider the masses of equal volumes of water, silver, and gold, as determined by means of a beam balance. The mass of a cubic centimeter of water is one gram (1 g); that of a cubic centimeter of silver, 10.5 g; gold, 19.3 g. That is to say, the density of water is 1 g per cm^3 (1 g/cm^3); the density of silver, 10.5 g/cm^3; gold, 19.3 g/cm^3. Density is sometimes expressed in pounds per cubic foot. Thus the density of water is 62.4 lb./cu. ft.; density of silver, $10.5 \times 62.4 = 655.2$ lb./cu. ft.; gold, $19.3 \times 62.4 = 1203.32$ lb./cu. ft.

It should be noted, however, that the densities of substances, as tabulated in Density Tables, are nearly always given in terms of grams per cubic centimeter (g/cm^3). The following is a list of the densities of a few common substances. For a more complete list of densities, see Table VII, Appendix.

DENSITY OF SOLIDS, IN G PER CM^3

Wood, pine	0.5	Copper	8.9
Ice	0.9	Nickel	8.9
Glass, crown	2.6	Silver	10.5
Zinc	7.1	Lead	11.3
Iron, gray cast	7.1	Mercury	13.6
Iron, wrought	7.8	Gold	19.3
Brass	8.5	Platinum	21.4

DENSITY OF LIQUIDS, G PER CM^3

Gasoline	0.75	Glycerine	1.26
Kerosene	0.8	Sulphuric acid	1.84
Water	1.0	Mercury	13.6

Since density is mass per unit of volume, it follows that

the density of a substance may be found by dividing its mass by its volume. This may be written

$$density = \frac{mass}{volume}$$

$$D = \frac{M}{V}$$

EXERCISES. **1.** (a) How many times heavier than water is an equal volume of glycerine? (b) What is the weight of a cubic foot of glycerine?

2. A gallon jug when full contains 8.3 lb. of water. (a) How many lb. of gasoline will the jug hold? (b) How many lb. of sulphuric acid?

3. A cylindrical vessel, $h = 20$ cm, $r = 4$ cm, is filled with a certain liquid, the mass of which is 743.931 g. (a) What is the density of the liquid? (b) Consult Table VII, Appendix, and find the liquid the density of which corresponds to this value. What is the liquid?

33. Force and Pressure. Force is expressed in terms of the push or pull upon the entire area of the surface affected. Pressure, on the other hand, always refers to the force exerted on a *unit* area. *Pressure* is force per unit area. That is,

$$pressure = \frac{force}{area}$$

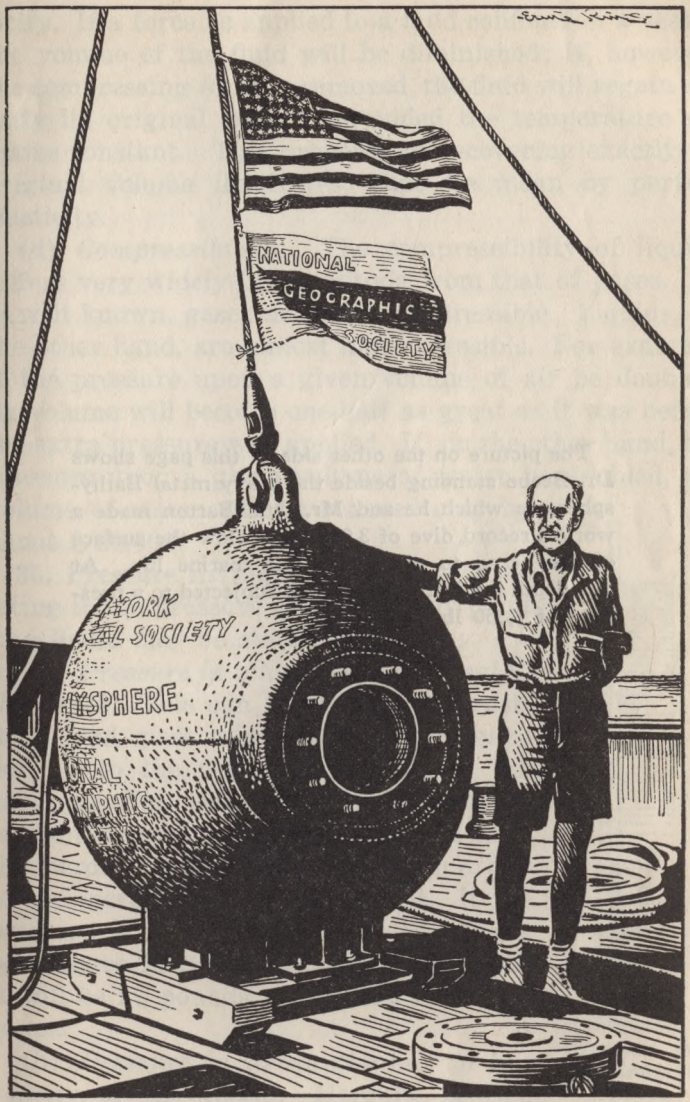
$$P = \frac{F}{A}$$

34. Properties of Fluids. Fluids possess certain general properties or characteristics which may be stated briefly as follows:

(a) *Pressure normal to surface.* The pressure which a fluid at rest exerts against the walls of the containing vessel is always normal (that is, at right angles) to the surface pressed upon.

(b) *Pressure at a point.* Pressure at any point in a fluid at rest is equal in all directions.

(c) *Elasticity perfect.* Fluids possess perfect elas-



READY FOR A 3,028 FOOT DIVE

The picture on the other side of this page shows Dr. Beebe standing beside the heavy metal Bathysphere in which he and Mr. Otis Barton made a world's record dive of 3,028 feet below the surface of the ocean to study deep-sea marine life. At this depth the Bathysphere was subjected to a pressure of 1,360 lb. per sq. in.

ticity. If a force be applied to a fluid confined in a vessel, the volume of the fluid will be diminished; if, however, the compressing force is removed, the fluid will regain exactly its original volume, provided the temperature remains constant. This property of recovering exactly its original volume illustrates what we mean by perfect elasticity.

(d) *Compressibility.* The compressibility of liquids differs very widely in magnitude from that of gases. As is well known, gases are very compressible. Liquids, on the other hand, are almost incompressible. For example, if the pressure upon a given volume of air be doubled, its volume will become one-half as great as it was before the extra pressure was applied. If, on the other hand, the pressure upon a given volume of water be doubled, the volume of water will be reduced only by about 1/20,000.

35. Pressure in Liquids. The facts relating to the pressure exerted at any point in a liquid may be stated as follows:

(a) *Pressure in a liquid proportional to the depth.* This can be shown by thrusting a bent glass tube containing mercury down into the water enclosed in a tall glass jar, Fig. 18. As the curved portion of the tube moves downward from *a* to *b*, the mercury is depressed in the arm *c*, due to the increased pressure upon it. If the pressure at a given depth be 10 pounds per square inch, then at twice the depth it will be 20 pounds per square inch, and so on.

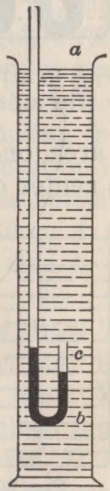


FIG. 18
Pressure proportional to depth

(b) *Pressure is proportional to the density of the liquid.* Mercury is 13.6 times as heavy as water; therefore, for a

given depth, mercury will exert 13.6 times as great a pressure as water.

(c) *Pressure at any point in a given plane parallel to the surface is equal in all directions.* This fact may be shown by thrusting three bent tubes containing mercury into a vessel of water, Fig. 19. The lower openings of the tubes are all in the same plane, but communicate with the water in different directions. The depression of the mercury in each tube is the same, showing that the pressure in all directions in the plane LL' is the same.

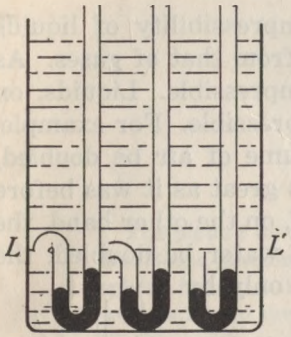


FIG. 19.—Pressure equal in all directions

36. Pressure at Any Depth in a Liquid. Let SS', Fig. 20, represent the surface line of a liquid in a vessel. Suppose that we wish to find the pressure at a point a below the surface, at the level LL'. Select an element of area, one square centimeter, say, in the level LL', with its center of area at the point a. Now all we have to do to find the pressure of the liquid at this point is to find the weight of a column of liquid of unit cross section and of a height equal to H, the vertical distance from a to the surface. Weight is volume times density, and volume is cross section area times length times density. But since pressure is force per unit area and the cross sectional area in question is 1, and hence may be omitted, we may therefore write

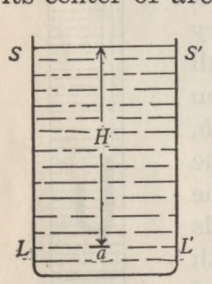


FIG. 20
Pressure = height of column × density

$$pressure = height\ of\ column \times density$$
$$P = HD$$

This equation is of great importance in Hydrostatics, and the student should get clearly in mind the meaning of the terms, P, H, and D. P is the pressure, that is, the force per unit area; H is not necessarily the depth of the liquid in the vessel but the *vertical distance from the surface of the liquid to the center of area of the surface pressed upon;* D is the density of the liquid in grams per cubic centimeter or in pounds per cubic foot, depending upon the units in which H is given. For example, if H is in centimeters, then the density D is in grams per cubic centimeter; if H is given in feet, then D is in pounds per cubic foot.

Example. Find the pressure in water when (a) $H =$ 10 cm; (b) $H = 10$ ft. *Solution*: (a) $P = HD = 10 \times 1 = 10$ g/cm^2; (b) $P = HD = 10 \times 62.4 = 624$ lb./sq. ft.

37. Force on the Bottom of a Vessel. Given a cylindrical tank having a cross sectional area of 6 square feet and a length of 10 feet, placed in an upright position, and filled with water, to find the force exerted by the water upon the bottom. Now there are two ways of finding the force on the bottom. First, since the vessel in question is of uniform cross sectional area, the force on the bottom is equal to the volume of the liquid ($6 \times 10 = 60$ cu. ft.) multiplied by its density (62.4 lb. per cu. ft.), that is

$$F = 60 \times 62.4 = 3744 \text{ lb.}$$

That is to say, the *force exerted on the bottom is equal to the weight of the liquid.* This is true, however, only in the case of vessels of uniform cross sectional area. It would not be true in the case of a vessel shaped like a jug, in which the cross sectional areas are not uniform. It is necessary, therefore, to find a second and more general solution of the problem.

The second method involves the determination of the force in terms of the area of the bottom and the *pressure* exerted upon it. Force is equal to area times pressure.

But we have just learned in Art. 36 that $P = H\ D$. It follows then that

$$force = area \times pressure$$
$$F = A\ H\ D$$
$$= 6 \times 10 \times 62.4 = 3744\ lb.$$

The equation $F = AHD$ applies to vessels of any shape whatsoever. In computing the force, all we have to know is the area of the bottom, the vertical distance from the surface of the liquid to the bottom, and the density of the liquid.

38. Force on the Side of a Vessel. By way of illustration of the method of finding the force exerted by a liquid on the side of a vessel, let us suppose that we have a rectangular tank 10 feet in height and 5 feet in width on each side. The tank is in an upright position and is filled with water. We wish to find the force exerted by the water on one side of the tank. The depth of the water is 10 feet and the area of one side is $5 \times 10 = 50$ square feet.

Now since a liquid is practically incompressible, the pressure exerted is proportional to the depth. The pressure at the surface is zero, and at the bottom it is a maximum. But what we want to know is not the pressure at the surface nor at the bottom, but the *average pressure.* The place of average pressure is at a point halfway between the surface and the bottom; that is, the length H in this case is *one-half* the depth of the liquid. The force on the side of the vessel then is equal to the area of the side multiplied by the average pressure, and since the average pressure is one-half the depth of the liquid times its density, we may write

$$force\ on\ side = area\ of\ side \times \tfrac{1}{2}\ depth\ of\ liquid \times density$$

In the case of the problem in question, one-half the depth = 5 ft., and the density = 62.4 lb. per cu. ft. Then

$$F = 50 \times 5 \times 62.4 = 15{,}600\ lb.$$

39. Force on Any Plane Submerged Surface. The equation $F = AHD$ is perfectly general in its application, whatever be the position or size of the submerged surface. It tells us that the force exerted upon any surface depends upon three factors and three only, namely: A the area, H the vertical distance from the surface of the liquid to the *center of area* of the surface pressed upon, and D the density of the liquid. Let us, as an example, take the following case: Suppose that a thin piece of metal be submerged to the level LL', Fig. 21. Suppose that the area of this rectangular body is placed successively in three different positions, *I*, *II*, and *III*, the *center of area of the figures always being kept in the line* LL'. In *I* the body lies with its surface parallel to the surface of the liquid; in *II* it is placed in a slant wise position; and in *III* it is vertical. Imagine that we are looking at the submerged rectangular figure side on. Now since the factors A, H, and D are the same in all three cases, it follows that the force is the same for all three positions.

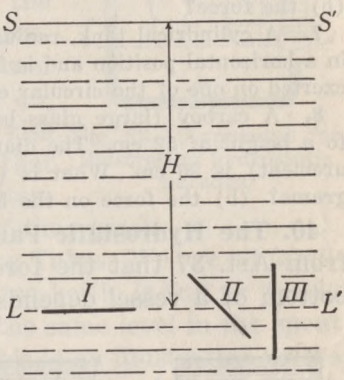


FIG. 21.—Equal forces on Plates *I*, *II*, and *III*

The *center of area* of a rectangle is at its middle point, that is, at the intersection of its diagonals; the center of area of a triangle is at the intersection of its median lines; that of a circle, at its center.

EXERCISES. **4.** A rectangular tank, the cross sectional dimensions of which are $5 \times 5 = 25$ sq. ft., and the length 15 ft., stands upright in a vertical position. It is filled with water. Find (a) the pressure on the bottom; (b) the force exerted on the bottom.

5. Consider again the tank described in Ex. 4. (a) What is the

pressure at the top? (b) at the bottom? (c) What is the average pressure? (d) What is the force exerted on one side?

6. A cylindrical tank has a cross sectional area of 28 sq. ft. and a length of 20 ft. The tank is placed in a vertical position and filled with gasoline. What is (a) the pressure on the bottom? (b) the force?

7. A cylindrical tank, radius 3 ft. and length 20 ft., is placed in a horizontal position and is filled with kerosene. Find the force exerted on one of the circular ends.

8. A carboy (large glass bottle) is filled with sulphuric acid to a height of 52 cm. The diameter of the bottom (inside measurement) is 30 cm. What is (a) the pressure on the bottom in grams? (b) the force on the bottom in kilograms?

40. The Hydrostatic Paradox. We have just learned from Art. 37 that the force exerted by a liquid on the bottom of a vessel depends only on the area of the bottom, the depth of the liquid, and its density. That is, the force is independent of the shape of the vessel. This is called the "hydrostatic paradox." If three vessels, *A*, *B*, *C*, Fig. 22, having bottoms of the same area, are filled to the same depth with a given liquid, the downward force exerted on the bottom of each vessel will be the same. In *A* the force on the bottom is equal to the weight of the liquid; in *B* the force is less than the weight of the liquid; in *C* the force is greater than the weight of the liquid. This can be demonstrated by means of three vessels, as shown in Fig. 23. The bottoms of the vessels have equal areas. The tube *T* contains mercury, which stands at the level *cd*. When vessel *A* is screwed upon the

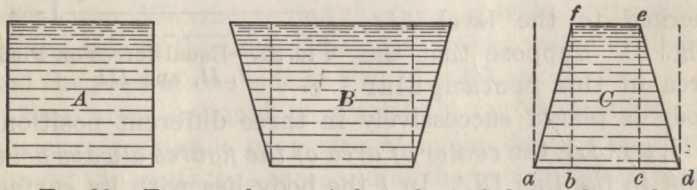


FIG. 22.—Force on bottom independent of shape of vessel

standard at d and is filled with water, the mercury falls in one arm and rises in the other from c to e, the height ce measuring the downward force at d. If now the vessels B and C be placed successively upon the standard d and filled with water to the level ab, the mercury in each case will rise to e as before, thus showing that the force exerted upon the bottom is independent of the shape of the vessel.

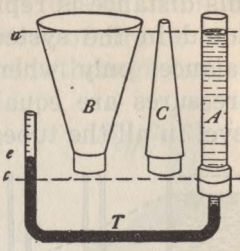


FIG. 23.—Apparatus for illustrating hydrostatic paradox

41. Liquids in Communicating Tubes. We are all familiar with the fact that when a tea kettle is full or nearly full, the water in it stands at the same level in the spout as in the body of the kettle. This is an illustration of the old saying, "Water always seeks its level." This principle may be demonstrated by means of the communicating tube apparatus, one form of which is shown in Fig. 24. When water is poured into one of the tubes, it rises to the same level in all of them, no matter what be their shape.

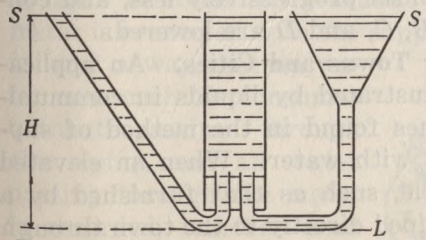


FIG. 24.—Communicating tubes, liquid at rest

An explanation of the rise of the water to a common level lies in the fact that the pressure in each tube is the same. We have learned that for a given liquid the pressure is dependent not at all upon the shape of the vessel but upon the depth of the liquid, the depth being measured in a vertical direction from the surface line of the liquid to a given level. In the apparatus shown in Fig. 24

this distance is represented by the vertical line *H*. The liquid in the system can be in equilibrium, that is in balance, only when the pressures are equal, and the pressures are equal when the liquid rises to the same level in all the tubes.

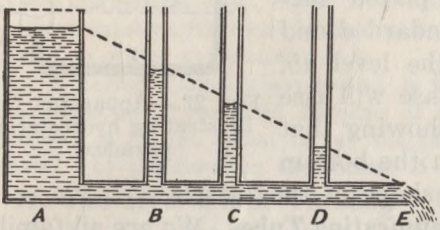


FIG. 25.—Communicating tubes, liquid flowing

The principle that liquids in communicating tubes seek a common level holds only for liquids at rest. If the liquid is in motion, the levels in the various tubes are not the same, Fig. 25. As we go from vessel *A* toward the opening *E* from which the water flows, the pressure becomes progressively less, and consequently the levels in *B*, *C*, and *D* are lowered.

42. Water Supply for Towns and Cities. An application of the principle illustrated by liquids in communicating tubes is sometimes found in the method of supplying towns and cities with water. When an elevated water supply is available, such as that furnished by a reservoir, the water is piped directly to the town through mains. When no such elevated supply is available, the water is pumped from a lake of lower level, from a river, or from wells into a stand pipe, Fig. 26, which is located upon the highest point adjacent to the town. From the stand pipe it is distributed to the residences and business places of the town. In cities of any considerable size, however, the stand pipe system is not in general use, because the supply thus furnished would be inadequate. Water is forced directly into the mains by high pressure pumps. The pressure exerted by the water in the mains is in general about 40 pounds per square inch.

Water in a modern city system is drawn from the mains through hydrants, as in the case of fire or when the streets are flushed, and in households through faucets.

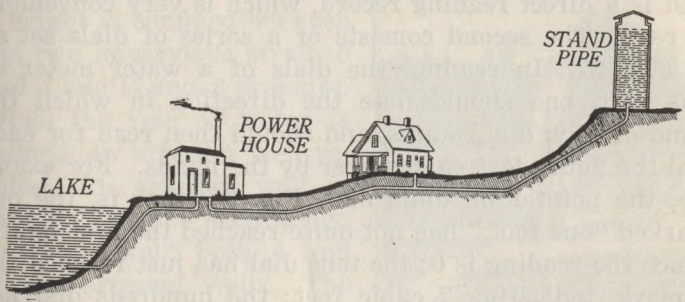


FIG. 26.—Water supply system showing power house and standpipe

43. The Water Meter. Water is usually sold at so many cents per thousand gallons and the quantity used by each customer is measured by a water meter. The meter is connected with the supply line in the water main in such a way that no water can reach the consumer without passing through the meter, and all water passing through the meter is recorded, whether used or wasted through leakage or otherwise.

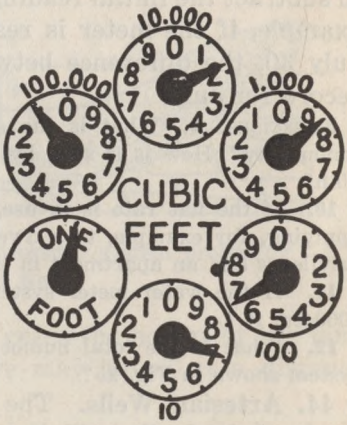


FIG. 27.—Recording dials of a water meter

There are several types of meters in use, the commonest being that known as the "disc valve" type. In this meter there is a disc-like valve which is supported at its center on a large ball and which tilts about on its edge as the water passes through the chamber.

The rotating valve is connected with a train of wheels which in turn move the recording dials. Water meters are provided with two types of recording system. The first is a direct reading record, which is very convenient to read. The second consists of a series of dials set as in Fig. 27. In reading the dials of a water meter of this sort, one should note the direction in which the hand of each dial rotates, and should then read for each dial the figure *last passed over* by the hands. For example, the pointer on units dial, Fig. 27, that is, the one marked "one foot," has not quite reached the first figure, hence the reading is 0; the tens dial has just reached the 7 mark, indicating 7 cubic feet; the hundreds dial lies between 6 and 7, and, therefore, should be read 60 cubic feet. The readings are then taken in consecutive order from the units dial to the 100,000 dial respectively.

In order to determine the amount of water which has passed through the meter in a given time, it is necessary to subtract the initial reading from the final reading. For example, if the meter is read on June 30, and again on July 30, the difference between the two readings is the record for July.

EXERCISES. **9.** What is the source of water supply for your community? How is it sold, by the flat rate or the water meter plan?

10. If the flat rate is in use, how is this rate determined, as applying, for example, to a dwelling house with lawn sprinkling privileges and an apartment in the downtown district?

11. If the water meter system prevails, what is the rate per 1000 cu. ft.?

12. What is the total number of cu. ft. recorded by the dial system shown in Fig. 25?

44. Artesian Wells. The tendency of water to seek its level is further illustrated by flowing, or artesian wells, Fig. 28. The ground conditions necessary for such a well are shown in Fig. 29. The first stratum is of loose soil; *b* is an impervious layer of clay or rock; *c*, a stratum

of gravel; and below this again is another impervious layer, *d*. Water flowing from a height is confined between the two impervious strata *b* and *d* and is under pressure. When the vein of water is tapped, an artesian well results. An artesian well is usually, but not necessarily, a flowing well; both *A* and *B* are artesian wells. The name "artesian" comes from the province of Artois, in France, where the operation of drilling for flowing wells was first performed.



FIG. 28.—Artesian well

Sometimes these flowing wells are found in compara-

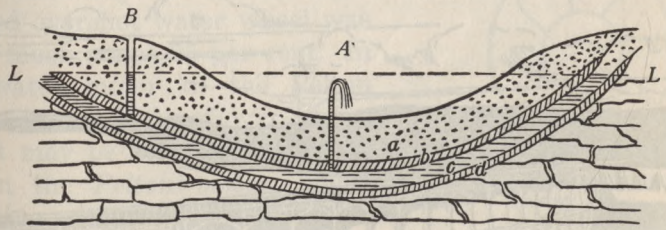


FIG. 29.—Illustrating the pressure system producing artesian wells

tively level regions. In such cases the height of the land which furnishes the pressure may lie off at a distance of a number of miles.

45. Water Wheels. One of the oldest methods of obtaining power from water was by means of the water wheel, one of the old-fashioned and picturesque types being shown in Fig. 30. In a device of this sort a

stream of water was allowed to flow over, or in some cases, under the wheel, thus causing it to revolve. It is needless to say that water wheels of this sort were very wasteful of power. They have long since been discarded for more efficient types, such as the Pelton water motor and the water turbine.

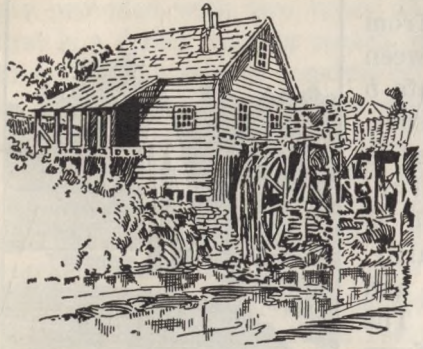


FIG. 30.—An old fashioned mill and water wheel

Modern water power plants, Fig. 31, have come to play an extremely important part in the industrial life of the nation. The hydroelectric systems of the

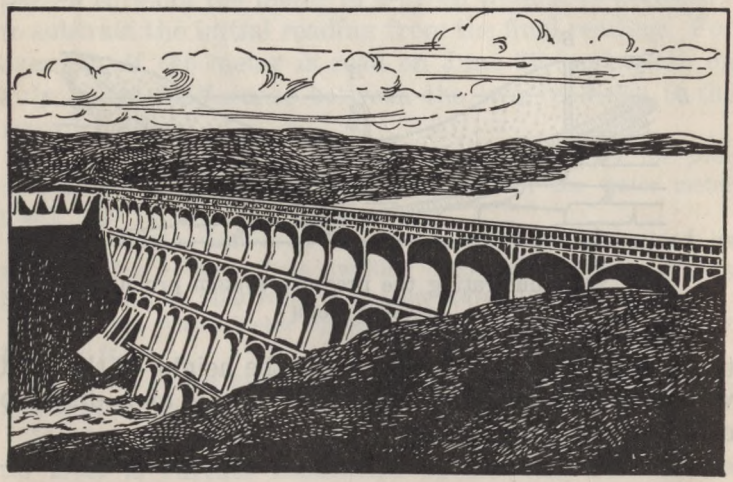


FIG. 31.—The Big Meadows Dam hydroelectric power plant

West constitute in the aggregate one of the biggest industrial developments this country has ever known. This de-

velopment is due not only to the demands for power, but to the fact that because of its topography this section of the country is particularly favorable for the generation of electric current on an extensive scale by means of hydroelectric plants. It is estimated by government experts that out of a total of about 26,000,000 horsepower which it is possible to develop in this way in the entire country, nearly 19,000,000 horsepower lies west of the Rocky Mountains. Conditions for hydroelectric development are particularly favorable in the three states of the Pacific coast where the waters from the melting snows, discharged down the precipitous slopes of the high sierras, provide one of the greatest natural resources in the world for the production of electrical energy.

46. The Water Motor. Modern water wheels are of two general types, namely, wheels of the Pelton type, Fig. 32, and those of the Turbine type. The efficiency of the old-fashioned overshot water wheel was not much above 25 per cent. In a water motor of the Pelton type, an efficiency of over 80 per cent may be attained.

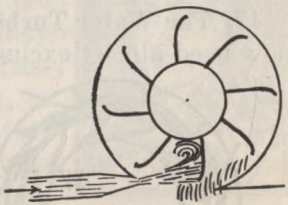


FIG. 32.—Pelton water motor

In the *Pelton motor,* water issuing with great velocity from a nozzle is directed against the curved blades of the wheel. The high efficiency of this type of water wheel is due to the curved blades which are employed. In order to explain the efficiency of a curved blade, as compared with a flat blade, let us consider the force which a jet of water exerts when it strikes against the blade of a water wheel. Suppose first that a jet of water be directed against a flat blade, Fig. 33. The force exerted in this case is proportional to the impact of the water. In the second place, let us consider what happens when the

same jet strikes a curved blade, as shown in Fig. 34. In this case the direction of the stream is completely re-

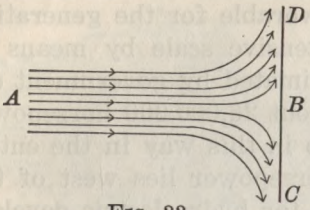


FIG. 33

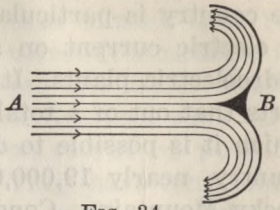


FIG. 34

Force of jet on curved blade twice as great as that on flat blade

versed, hence the force exerted is not only that due to the impact of the stream but also that required to throw the stream back. The force which a jet of water exerts upon a curved blade, as shown in Fig. 34, is theoretically twice as great as that which the same stream would exert upon a flat blade.

47. The Water Turbine. The water turbine is the type now used almost exclusively in large water power plants.

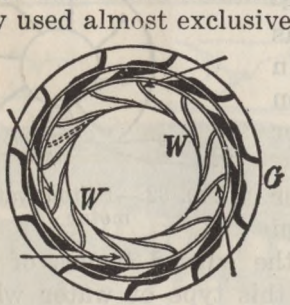


FIG. 35

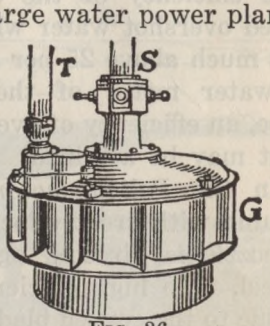


FIG. 36

Turbine water wheel

This wheel consists of a system of curved blades, Fig. 35, which are set to rotate in a turbine case, Fig. 36. The amount of water entering the wheel is regulated by the shaft *T*, by means of which the valves or gates *G* at the side are opened.

In modern hydroelectric plants the turbine is placed

at the bottom of a turbine pit, the bottom of which is as far below the surface of the water of the dam as possible. The wheel is enclosed in a turbine case which communicates through an intake pipe with a source of supply. The water enters at the top or side of the wheel, depending upon the type used, and comes out at the bottom. One of the largest turbines in existence is that operated by the Puget Sound Power Company. This plant develops 25,000 horsepower under a head of 444 feet of water.

Pascal's Law and Its Application

48. Pascal's Law. Pressure applied to a fluid confined in a vessel is transmitted equally in all directions to all parts of the fluid. The law relating to the transmission of fluid pressure was first definitely announced by Pascal (1623-1662), a French physicist. *Pascal's law* may be stated as follows: *Pressure applied to a given area of a fluid enclosed in a vessel is transmitted undiminished to every equal area of the vessel.* Pascal performed many experiments to demonstrate the application of this law. One of the most striking was his tube and cask experiment. A long tube was firmly fixed into the head of a stout cask, Fig. 37, and then both the cask and the tube were filled with water. The weight of the water in the tube exerted a force on the cask as



Pascal



Fig. 37

many times greater than itself as the inner area of the cask was greater than the cross sectional area of the tube. The force exerted by the water in the tube was sufficient to burst the cask. The student should note that only a part of the tube in question is shown in Fig. 37. Pascal found it necessary to use a relatively long tube.

EXERCISES. **13.** A jug which is filled with water, and into the neck of which there is inserted a tightly fitting stopper, may easily be broken by striking the stopper a sharp blow. The cross sectional area of the stopper in a given jug, which is full of water, is 1 sq. in. The inside area of the bottom of the jug is 100 sq. in. If a force of 10 lb. be applied to the stopper, what will be the bursting force exerted on the bottom of the jug?

14. Suppose that the cross sectional area of the neck of the jug is only ¼ of a sq. in., what force will be exerted on the bottom by a blow of 10 lb. applied to the stopper?

15. Consider an experiment similar to the one which Pascal is said to have performed with the tube and cask. How long would the tube have to be in order to furnish a pressure of 40 lb. / sq. in.?

49. The Relation of Area to the Force Exerted. In order to get more clearly in mind the significance of Pascal's law as applied in the operation of modern hydrostatic lifting machines, let us consider Fig. 38. Here we have a sectional view of two cylinders connected by a tube, the system (cylinders and connecting tube) being filled with a liquid. In the cylinders there are accurately fitting pistons, P and P'. Now since according to Pascal's law pressure is transmitted to every like area it follows that a force F applied to the small cylinder P will exert an upward force on the large cylinder P' which will be as many times greater than the force applied as the cross sectional area of the large cylinder is greater than the

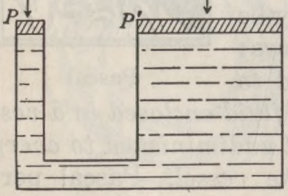


FIG. 38.—Illustrating Pascal's law

cross sectional area of the small one. Suppose that the diameter of the small piston is 1 inch and that of the large piston 1 foot. Suppose also that a force of 10 pounds is exerted on the liquid by the small piston P. What will be the lifting force exerted on the large piston? Since we know the magnitude of the force applied to the small piston, the question that concerns us is, How many times greater is the area of the large piston than that of the small one? The diameter of $P = 1$ in.; that of $P' = 1$ ft. $= 12$ in. Now the areas of the two cylindrical figures are to each other as the squares of their respective diameters. Hence the area of the large piston is to that of the small one as 144 is to 1; that is, the area of the large piston is 144 times as great as that of the small one. The upward force on the large piston is therefore $10 \times 144 = 1440$ lbs. This means that by such a device a child by exerting a force of 10 lbs. could lift a weight of 1440 lbs. We have here evidently a most remarkable device for the multiplication of force by the transmission of pressure.

50. Relation of Area to Displacement. In studying the principles involved in the application of Pascal's law, students not infrequently get the idea that we have a means of getting something for nothing; that is, that a small amount of work done on one piston will produce a large amount of work on the other. This is not true; indeed, less useful work is always done by the large piston than is expended upon the small piston, because of frictional and other energy losses. There is a gain in the *force* exerted but no gain in the work done. In order to understand this let us again study Fig. 38.

We have learned in the very beginning of our study, Art. 1, that energy is the capacity for doing work, and work is the overcoming of resistance through space. Work, then, when used in a physical sense, involves two

factors, namely, resistance (force) and space (displacement). When the small piston, Fig. 38, is pushed down, a definite amount of work is done, because force has been exerted through space; and likewise, when the large piston moves upward, a given amount of work is done. Now if it were not for energy losses due to friction and heat, these two amounts of work would be equal to each other. They are never quite equal, however, because there is always some loss.

We have learned also in the preceding article that the forces exerted by the two pistons are not equal; one may be very much greater than the other, depending upon their relative cross sectional areas. What we want to consider now is the relative spaces (displacements) through which the two pistons move. The small piston of the device shown in Fig. 38 has a cross sectional area of 1 square inch and that of the large piston 144 square inches. Now when the small piston is pushed down 1 inch, thus forcing 1 cubic inch of the liquid into the large cylinder, what fraction of an inch does the large piston rise? Clearly 1/144 of an inch. This means that if the large piston moves up 1 inch the small piston will have to be moved down 144 inches, or what amounts to the same thing, the small piston will have to move down 1 inch 144 times. There is then no gain in work. We have multiplied the force applied by 144, but in order to do this we had to move the small piston 144 times as far as the large piston moved.

We may summarize the above statements now by saying: (a) *The forces exerted by the pistons are directly proportional to their cross sectional areas;* and (b) *the spaces through which the pistons move are inversely proportional to their cross sectional areas.*

51. The Hydraulic Press. The most widely useful application of the principle stated by Pascal's law is found

in the operation of the hydraulic press, or as it is sometimes called, the hydrostatic press. This machine is used when it is necessary to exert enormous force, such as in the compression of material into a very small space for economy in shipment, for the lifting of heavy locomotives that have jumped the track, for the testing of building material, and for pressing heavy metal plates into shape.

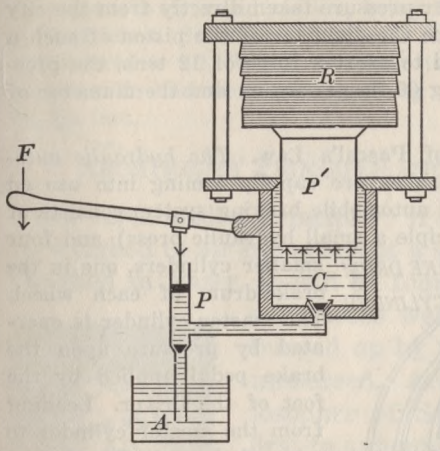


FIG. 39.—Sectional view of a hydraulic press

A sectional outline of a hydraulic press is shown in Fig. 39. The tank A contains a liquid which is driven by the piston P into the cylinder C, thus acting on the piston P', driving it upward. The force exerted upon the large piston P' is as many times greater than that exerted upon the liquid by the small piston as the cross sectional area of P' is greater than that of P.

EXERCISES. **16.** Two cylinders fitted with pistons, as shown in Fig. 38, and connected by means of a communicating tube, are filled with a liquid. This constitutes a simple form of the hydrostatic press. The cross sectional area of the small piston is 2 sq. in.; that of the large piston, 200 sq. in. A force of 20 lb. applied to the small piston will exert what upward force on the large piston?

17. When the small piston (Ex. 16) moves down 1 in., through what distance will the large one rise?

18. A weight of a ton placed on the large piston (Ex. 16) will exert what upward force on the small piston?

19. When the large piston (Ex. 16) moves down 1 in., through

what distance will the small piston move upward, assuming that the tube is long enough to permit of its upward motion?

20. Hydraulic presses are often designed for use in shops and factories to operate on water pressure taken directly from the city mains. What will have to be the diameter of the piston of such a press, provided it is desired to exert a force of 12 tons, the pressure in the city mains being 60 lb. per sq. in. and the diameter of the intake pipe 1 inch?

52. Other Applications of Pascal's Law. *The hydraulic automobile brake.* Hydraulic brakes are rapidly coming into use on automobiles. The hydraulic automobile braking system consists of a master cylinder (in principle a small hydraulic press) and four smaller cylinders, one in the brake-drum of each wheel. The master cylinder is operated by pressure upon the brake pedal applied by the foot of the driver. Leading from the master cylinder to each of the cylinders in the brake-drums are metal tubes and heavy non-expanding hose. In the master cylinder there is one piston, while in each of the brake-drum cylinders there are two pistons which operate from the center of the cylinder outward. These brake-drum pistons are connected to the two sections of the brake shoe, Fig. 40. The whole system, that is the master cylinder and brake-drum cylinders and connecting lines, are filled with a non-freezing and non-corrosive fluid. The pistons are held in place by a spring so that when there is no pressure upon the brake pedal, contact between the brake-shoe and the drum is broken, that is the brake is "off." When the brake pedal is depressed, the piston in the master cylinder moves forward, expelling into the lines sufficient liquid to force out the pistons in each of the brake-drum cylinders until the brake-shoes come in contact with the drums; in other words, the brakes are "on".

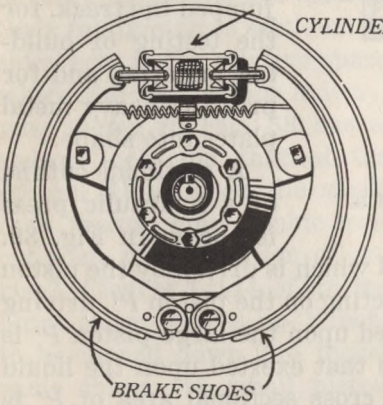


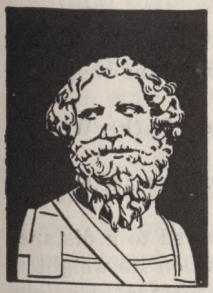
FIG. 40.—Hydraulic automobile brake

The hydraulic chair. The hydraulic chair is now used generally

in barber shops and dental offices. The principle upon which such a chair works is essentially that of the hydraulic press. In this case, however, the piston upon which the chair rests moves with a comparatively rapid motion, being forced upward by means of a hand lever. Sometimes this lever is operated by the foot. The liquid used is oil. When the chair is lowered, the oil from the compression chamber is released by means of a lever operated by the foot.

ARCHIMEDES' PRINCIPLE AND ITS APPLICATION

53. Archimedes' Principle. If a stone or piece of metal be weighed in air by means of a spring balance, and then weighed in water, it will be found that the body apparently loses weight in the liquid. It is buoyed up by the liquid in which it is immersed. Archimedes, a Greek philosopher of Syracuse, Sicily, was the first to announce the principle of buoyancy, which has come to be called *Archimedes' principle,* and which may be stated thus: *A body immersed in a fluid is buoyed up by a force equal to the weight of the fluid displaced.*



Archimedes

That the above statement is true may be demonstrated by the following experiment. Given a metal cylinder *W* to find how much it is buoyed up in water; that is, to find its apparent loss of weight in water. Fill an overflow vessel *A*, Fig. 41, with water, up to the level of the outlet. Weigh a dry, empty beaker and place it under the spout of the overflow vessel. Now weight *W* in

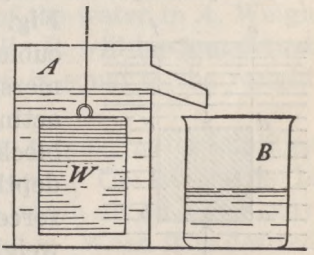


FIG. 41.—A body immersed in a fluid is buoyed up by a force equal to the weight of the fluid displaced

air and then in water, noting its loss of weight. Also find the weight of the water which has overflowed into B (weight of beaker and water — weight of empty beaker). The volume of the water in B is equal to the volume of W. Now it will be found that the loss of weight of W (weight in air—weight in water) is just equal to the weight of the water in B. Hence we may say that W is buoyed up by a force equal to the weight of the water displaced.

By way of summary, the student should note two things: first, *a body completely immersed in a fluid displaces its own volume; second, the loss of weight is equal to the weight of the fluid displaced.*

EXERCISES. **21.** Suppose that the weight of W, Fig. 41, is 340 g in air, and that it displaces 40 g of water. What is the weight of W in water?

22. (a) What is the volume of a gram of water? (b) What is the volume of W (Ex. 21) in cm^3?

23. A stone having a volume of 1 cu. ft. weighs 168.6 lb. in air. What will it weigh in water?

24. By means of a string, an iron ball is lowered to a distance of 1 ft. below the surface of water. In this position its weight is 8 lb. What will be its weight 10 ft. below the surface? Explain.

54. The Principle Explained. An explanation of Archimedes' principle of buoyancy may be given by means of Fig. 42. Let *a b c d* be one face of a cubical block of unit thickness immersed in water; the downward force acting upon the upper face of the block is equal to the area *cd* times the depth of *cf;* that is, the downward force on the block is equal to the weight of the column of liquid *d c f e*. The upward force acting on the lower face of the block, due to the liquid, is equal to the area of *ab* times the depth of the liquid *bf;*

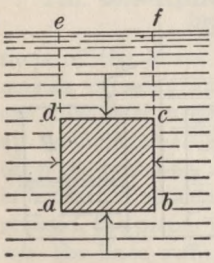


FIG. 42

that is, the block is buoyed up by the weight of a column of liquid proportional to *a b f e*. The upward force is therefore greater than the downward force, and the difference between the two, (*a b f e*)—(*d c f e*), is equal to the weight of a column of liquid *a b c d;* hence we say that the block is buoyed up by the weight of the water which it displaces.

Whether a body immersed in a fluid will sink or float depends upon the relation between its own weight and the weight of the fluid displaced. If a body is heavier than the fluid displaced, it will sink; if lighter, it will float. If when completely immersed it weighs just as much as the fluid displaced, it will tend neither to sink nor float but will remain stâtionary in the fluid wherever placed.

55. The Principle of Flotation. Archimedes' principle applies equally well to bodies that float as to bodies that sink. A body that sinks displaces its own volume; a body that floats displaces its own weight. We may again use the overflow apparatus to prove this. Given a piece of wood, or other body lighter than water, to show that when placed upon water it will displace an amount of the liquid equal to its own weight. Fill the overflow vessel, Fig. 41, as before. Dry and weigh the beaker *B* and place it under the spout. Weigh the wood in air, and then place it carefully on the surface of the water in *A*. Weigh the water which overflows into *B*. It will be found that the weight of the water displaced is equal to the weight of the wood.

The navigation of the modern submarine exemplifies beautifully this principle of buoyancy. The submarine is provided with ballast tanks, into which air or water may be forced at the will of the navigator. When it is desired to submerge the vessel, water is allowed to enter the ballast tanks until the density of the craft becomes nearly equal to that of sea water. The boat is then almost

entirely submerged. In order now to submerge it completely to a given level below the surface, the vessel is put under motion and the diving rudders are tilted downward; when it is desired to bring the vessel to the surface, the diving rudders are tilted upward. When it is necessary to bring the vessel to the surface for supplies or for other purposes, the water in the ballast tanks is forced out by air from the compressed air tanks, thus reducing the weight of the vessel. When running just below the surface, objects on the surface are viewed through a periscope, which contains a system of reflecting mirrors. On the surface, submarines are propelled by means of oil engines; below the surface by electric motors driven by storage batteries.

EXERCISES. **25.** A piece of paraffin weighs 126 g in air. It is placed successively on water and then on mercury. (a) How many grams of water does it displace? (b) How many grams of mercury?

26. The paraffin (Ex. 25) is stuck on the end of a stiff piece of fine wire and is thrust below the surface of water. It displaces 140 g of water. What is (a) the volume of the paraffin? (b) its density?

27. A row boat weighs 120 lb. A man weighing 160 lb. gets into the boat. What weight of (a) fresh water will the two displace? (b) sea water?

28. A steam boat sails down the Hudson River into the Atlantic Ocean. When it reaches the ocean, will it float higher or lower in the water? Explain.

29. The displacement of ships is expressed in tons. For example, the displacement of the "Berengaria," one of the largest ocean-going liners, is said to be 57,000 tons. This means that that portion of the ship which is below the surface displaces 57,000 tons of water. (a) What is the weight of the ship? (b) How many cu. ft. does this displacement represent, the weight of a cu. ft. of sea water being 64.3 lb.?

56. Density and Buoyancy. Many interesting experiments may be performed to illustrate the relation between density and buoyancy. In reading what follows,

the student should bear in mind the following facts which he has already learned: (a) *A body that floats displaces its own weight of the fluid.* (b) *A body that sinks displaces its own volume.* (c) *In both cases, that is, floating or sinking, the body is buoyed up by a force equal to the weight of the fluid displaced.*

Iron on mercury. The density of mercury, expressed in grams per cubic centimeter, is 13.6 and that of cast iron 7.1. A piece of iron, then, when placed on mercury will float. When some metals, such as zinc and brass, come in contact with mercury an amalgam is formed, that is to say, the mercury penetrates into the metal. Iron is not thus affected by mercury. This is the reason for using iron in this experiment. It would not do to attempt to float zinc or brass on a mercury.

Iron on water. Practically all modern ships are made of iron. Everyone knows that iron is heavier than water. How then does it happen that an iron ship floats? If a piece of iron in the form of a solid mass is placed in water it will at once sink. If, however, the iron be cast in the form of a hollow vessel and placed in water with the open side up, it will float, because the combined weights of the vessel and the air which it contains is less than the weight of an equal volume of water. If the vessel weighs 800 grams it will displace 800 grams of water when placed in a floating position; that is to say, when the vessel is floating in an upright position on water, it will sink until it displaces 800 grams of water. Of course, if the iron vessel be tipped over so that the air is displaced by water, it will sink.

Solids suspended in liquids. Expressed in grams per cubic centimeter the density of mercury is 13.6; cast iron, 7.1; water, 1; paraffin, 0.9; kerosene, 0.8; and pine wood, 0.5. A striking experiment to illustrate the relation between density and buoyancy in the case of the substances

mentioned above may be performed as follows: Pour into a tall cylindrical vessel some mercury, then about the same volume of water, and finally about the same volume of kerosene. There are now three layers of liquid in the vessel, namely, mercury, water, and kerosene. Now drop into the vessel an iron ball, a paraffin ball, and a wooden ball. The iron ball floats on the surface of the mercury; the paraffin floats on the surface of the water; and the wood floats on the kerosene, as shown in Fig. 43.

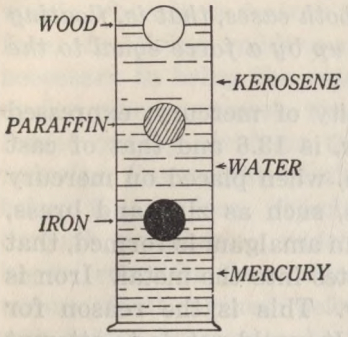


FIG. 43

SPECIFIC GRAVITY AND DENSITY

57. Specific Gravity. If in referring to a piece of iron we say, "The iron is heavy," we make a statement that is more or less indefinite as to meaning, because we may imply that the iron is heavy in the sense that considerable muscular effort has to be put forth in order to lift it, or we may mean that the piece of iron in question is heavy as compared with some other body, as a block of wood or a stone. The statement, "Iron is heavy" is more definite than the one given above, because here we are manifestly comparing the weight of iron as a substance with that of some other substance. In order, however, to make such a comparison specific, we must choose some substance as a standard of comparison, and furthermore, we must compare equal volumes. If, for example, we say that cast iron is 7.1 times as heavy as an equal volume of water, we are making a very definite statement indeed.

Now as commercial relations developed between men and nations, it became necessary to compare the weights of different substances one with the other. To do this a standard of comparison became necessary. Naturally the standard chosen was water, one of the most commonly occurring substances in nature. And so there grew up the idea of "specific weight" which later came to be called "specific gravity," since it became known that the weight of a body is the measure of the force of gravity acting upon it.

Specific gravity is the ratio of the weight of a body to the weight of an equal volume of water. When it is desired to make the determination very accurate, distilled water at 0° C is employed. In ordinary determinations, however, water at room temperature is used.

58. **Relation of Specific Gravity to Density.** The specific gravity of a substance is the ratio of the weight of a given portion of the substance to that of an equal volume of water. It follows then that in the determination of specific gravity we have simply to divide the weight of the given body by the weight of an equal volume of water. This is a general statement of the principle involved.

Density, on the other hand, is the mass of a body divided by its volume; that is, $D = M/V$.

What we are interested in here is to determine the numerical relation which exists between the *specific gravity* units and the *density* units for a given substance. In order to determine this relationship, as expressed in English and metric units, let us compare the specific gravity of lead with its density. The mass of a cubic foot of lead, and therefore its weight at sea level, is 705.12 pounds. The weight of a cubic foot of water is 62.4

pounds. In *English units* then, both for specific gravity and density we have

$$\textit{specific gravity} = \frac{\textit{weight of body}}{\textit{weight of equal volume of water}}$$

$$= \frac{705.12 \text{ lb.}}{62.4 \text{ lb.}} = 11.3$$

$$\textit{density} = \frac{\textit{mass}}{\textit{volume}}$$

$$= \frac{705.12 \text{ lb.}}{1 \text{ cu. ft.}} = 705.12 \text{ lb./cu. ft.}$$

Thus it appears that specific gravity (11.3) and density (705.12 lb./cu. ft.) are *not* numerically equal to each other, *when expressed in English units.*

Let us now make a comparison of the specific gravity of lead with its density, as expressed in *metric units.* The mass of 5 cubic centimeters of lead is 56.5 grams. The weight of an equal volume of water is 5 grams. We may therefore write

$$\textit{specific gravity} = \frac{56.5 \text{ g}}{5 \text{ g}} = 11.3$$

$$\textit{density} = \frac{56.5 \text{ g}}{5 \text{ cm}^3} = 11.3 \text{ g/cm}^3$$

Specific gravity (11.3) and density (11.3 g/cm^3) *are* numerically equal to each other, *when expressed in metric units.* It is important to keep this in mind, because density tables are now almost universally given in g/cm^3. Thus, when we find in a density table (Table VII, Appendix) that the density of the rare metal osmium, for example is 22.5 g/cm^3, we know at once that its specific gravity is also 22.5. Likewise, when we state the specific gravity of a substance (gold) is 19.3, we know that its density is 19.3 g/cm^3. Specific gravity is a ratio, an abstract number. Density, on the other hand, is not an abstract number; it is mass per unit volume.

EXERCISES. **30.** Consult Table VII, Appendix, and find the name of the metal referred to above, the density of which is 19.6 g/cm³.

31. Which has the greater specific gravity, tungsten or gold?

32. A given volume of a certain metal weighs 12.67 lb. The weight of an equal volume of water is 1.2 lb. (a) What is the specific gravity of the metal? (b) What is its density? (c) What is the metal?

59. Methods of Determining Specific Gravity and Density. *Solids heavier than water.* Specific gravity is the ratio of the weight of a body to the weight of an equal volume of water. But according to Archimedes' principle the weight of an equal volume of water is equal to the difference of the weight of the body in air and its weight in water. The weight in air may be determined by placing the body on the scale pan of a beam balance, and its weight in water by suspending it from the scale pan, Fig. 44. When these two weights have been found and the loss of weight determined, to find the specific gravity we have only to divide the weight in air by the loss in water, that is,

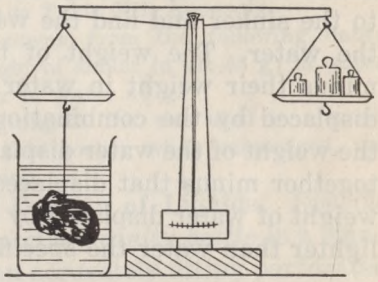


FIG. 44.—A method of determining the density of a solid

$$\textit{specific gravity} = \frac{\textit{weight in air}}{\textit{loss of weight in water}}$$

Example. The weight of a piece of glass in air is 37.7 g; its weight in water, 23.2 g; loss of weight = 37.7 — 23.2 = 14.5 g. Then since we are dealing with metric units

$$\textit{specific gravity} = \frac{37.7}{14.5} = 2.6$$

$$\textit{density} = 2.6\ g/cm^3$$

Solids lighter than water. To find the specific gravity

of a body which is lighter than water it is necessary to attach a sinker to the body. Naturally the sinker must be heavy enough to submerge the body. Sinkers used for this purpose are usually provided with a sharp pointer or needle upon which the body may be stuck. Let it be required to find the specific gravity of a body lighter than water, such as a piece of wood, beeswax, or paraffin. The method of procedure is as follows: First find the weight of the given substance in air. Next find the weight of the sinker in air and then in water. The difference between these last named weights is the weight of the water displaced by the sinker. Now fasten the body to the sinker and find the weight of the two together in the water. The weight of the body and sinker in air minus their weight in water is the weight of the water displaced by the combination (body and sinker). Now the weight of the water displaced by both body and sinker together minus that displaced by the sinker alone is the weight of water displaced by the body. Hence for a body lighter than water the specific gravity is

$$s.g. = \frac{\textit{weight of body in air}}{\textit{loss of weight body and sinker—loss of weight sinker}}$$

Example. To find the specific gravity and density of a given sample of wood from the following data: A piece of wood weighing 10 g is attached to a sinker which weighs 33 g in air and 30 g in water. The wood and sinker in water weigh 20 g. In order to determine the specific gravity of the wood, we must find three things, namely, (a) weight of wood in air, (b) loss of weight of sinker in water, and (c) the loss of weight of wood and sinker.

(a) Weight of wood in air = 10 g
Weight of sinker in air = 33 g
Weight of sinker in water = 30 g

(b) Loss of weight, sinker = 3 g
Weight wood and sinker in air = 43 g
Weight wood and sinker in water = 20 g

(c) Loss of weight of wood and sinker
$= 43 - 20 = 23$ g

$$\textit{specific gravity} = \frac{10}{23 - 3} = 0.5$$

$$\textit{density} = 0.5\ g/cm^3$$

EXERCISES. **33.** A piece of lead pipe was found to have a weight of 14.7 lb. in air, and 13.4 lb. in water. (a) What is the specific gravity of lead? (b) its density?

34. Find the specific gravity of copper from the following data: Volume of the copper, 1 cu. ft.; weight in air, 555.36 lb. Check your answer by referring to Table VII, Appendix.

35. Find the density of beeswax from the following data: Weight of wax in air 24 g; weight of sinker in air 46 g; in water 41 g; weight of sinker and wax in water 40 g.

36. A cast iron anchor weighing 20 lbs. was used in a small boat. What is the weight of this anchor when submerged, the specific gravity of cast iron being 7.4?

60. Specific Gravity and Density of Liquids. *Specific gravity bottle method.* A *specific gravity bottle* is a small glass vessel provided with an accurately fitting perforated glass stopper, Fig. 45. One of the most accurate methods of finding the specific gravity of a liquid is that known as the specific gravity bottle method. This method is in principle very simple. First find the weight of the empty bottle. Second find the weight of the bottle filled with the liquid whose specific gravity is to be determined. When the stopper is inserted, a small quantity of liquid will be forced out. This should be carefully wiped away before the weighing is made. The weight thus obtained minus the weight of the empty bottle is the weight of the liquid. Next find the weight of the bottle filled with distilled water, and subtract as before the weight of the bottle. We



FIG. 45
Specific gravity bottle

now have the weight of the given volume of the liquid and the weight of an equal volume of water. The specific gravity is therefore the weight of the liquid, divided by the weight of the water, that is

$$specific\ gravity = \frac{weight\ of\ liquid}{weight\ of\ water}$$

Sinker method. In laboratory practice it is sometimes desirable to find the specific gravity of a liquid by the sinker method. In this case a sinker of convenient size is weighed in air and then in the given liquid, after which it is weighed in water. The loss of weight in each case is that of a volume of the liquid equal to the volume of the sinker. That is to say, the two weights thus obtained are those of equal volumes of the liquid in question and the water. The loss of weight of the sinker in the given liquid divided by the loss of weight in water is the ratio required. We then may write

$$specific\ gravity = \frac{loss\ of\ weight\ of\ sinker\ in\ given\ liquid}{loss\ of\ weight\ of\ sinker\ in\ water}$$

61. The Hydrometer. A *hydrometer* is an instrument for determining the specific gravity of a liquid by placing the instrument in the liquid. It consists of a hollow sealed glass tube, which is weighted at one end so as to float vertically in the liquid, Fig. 46. The upper end consists of a narrow stem which is provided with a specific gravity scale, and the lower end is weighted with fine shot or mercury.

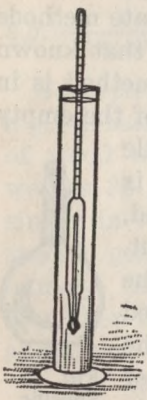


FIG. 46
Hydrometer

The principle by which the hydrometer is calibrated may be explained by the aid of Fig. 47. Three tall glass jars are filled nearly to the top with liquids of known specific gravities, number *I* containing water, specific gravity 1; *II* containing kerosene, spe-

cific gravity 0.8; *III*, glycerine, specific gravity 1.26. It will be noted that one of the liquids chosen has a specific gravity that is less than that of water and the other a specific gravity which is greater. A thin rod of hard wood of uniform cross section and of suitable length is weighted at one end with lead, so that when placed in the jar containing water, it floats in a vertical position about as shown. When the rod comes to rest in the water, a mark is made upon its side coincident with the surface of the water. This mark corresponds on the scale, which you are about to make, to the specific gravity of water, namely, 1. The rod is next placed in the jar containing kerosene. Since this liquid is less dense than water, the rod sinks to a lower level. A record mark is made on the rod at the surface of the kerosene to indicate the specific gravity of kerosene, that is 0.8. Finally the rod is placed in the jar containing glycerine. This liquid is more dense than water and consequently the rod floats at a higher level. The mark on the rod in this case indicates a specific gravity of 1.26.

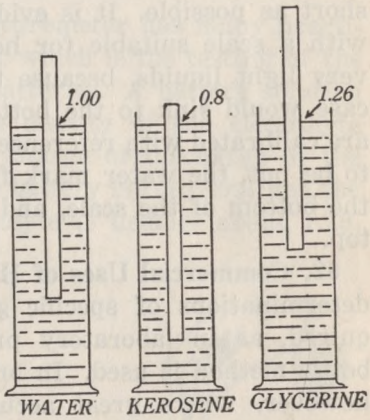


FIG. 47.—Calibrating a hydrometer

We now have three known points on our scale, representing the specific gravities of water, glycerine, and kerosene respectively. The space on the scale between the glycerine mark and the kerosene mark therefore represents $1.26 - 0.8 = 0.46$ specific gravity units. If now

we measure the linear distance between the kerosene and glycerine density marks in millimeters and divide 0.46 by the number of millimeters thus found, we shall have the value of one millimeter unit of linear distance in terms of specific gravity units. With this information available we may now extend the specific gravity scale above and below the water density mark within any desired limits.

For convenience in handling, hydrometers are made as short as possible. It is evident then that a hydrometer with a scale suitable for heavy liquids will not do for very light liquids, because the instrument in the latter case would sink to the bottom. Hydrometers therefore are calibrated with reference to the use to which they are to be put, the water mark for a light liquid being near the bottom of the scale, and for a heavy liquid near the top.

62. Commercial Uses of Hydrometers. When accurate determinations of specific gravity and density are required, as in laboratory practice, the specific gravity bottle method is used. In ordinary commercial practice, however, where great accuracy is not necessary, the hydrometer method is now almost universally employed, because of the rapidity with which such determinations may be made. The hydrometer is therefore found in daily use in industrial plants in the testing of acids, in the separation of different oils which are obtained from the distillation of crude petroleum, in the determination of the relative purity of vegetable oils, and in the grading of syrups in sugar factories. Milk inspectors use the hydrometer in testing milk for water adulteration. The hydrometer used for this purpose is called a *lactometer*. The specific gravity of cows' milk varies from 1.027 to 1.035. Since only the last two figures on the scale readings are significant, the numbers on the scale of a lactom-

eter usually run from 20 to 40, which means from 1.020 to 1.040. For example, if the reading of a given sample of milk is 35, this means that the specific gravity is 1.035. The student should note that such a test does not necessarily give accurate information as to the purity of the milk with reference to the presence or absence of dirt and disease germs; it gives us information mainly as to its water and butter fat content. Milk is about 87 per cent water.

It thus appears that the hydrometer has many practical uses, the most common of which is the testing of the acid solution of automobile batteries. A battery solution is called an *electrolyte*. The density of the electrolyte depends upon the degree of charge or discharge of the battery. When a lead storage battery is fully charged the specific gravity of its electrolyte is usually about 1.29, depending somewhat upon the make of the battery and the use for which it was designed. As the battery is discharged the density of the electrolyte decreases. A specific gravity of 1.15 indicates a low charge; it indicates that the battery needs to be recharged. It is important then that batteries be tested frequently. The instrument generally used for this pur-

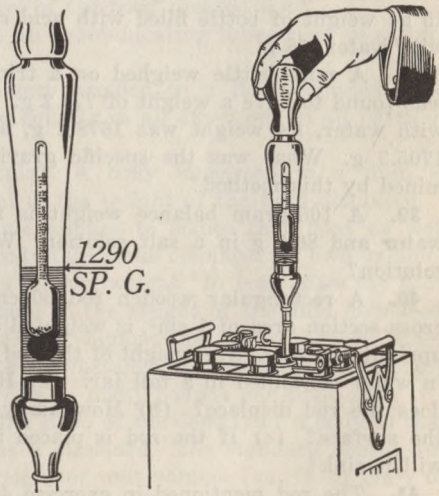


FIG. 48.—Battery hydrometer

pose is a hydrometer of the syringe type, Fig. 48. It looks exactly like a big medicine dropper, with a small hydrometer inside. To the lower end of the instrument there is attached a short rubber tube, and to the upper end a rubber bulb. When it is desired to make a specific gravity test, the tube is inserted into the solution through an opening in the top of the battery, and the bulb is compressed, thus forcing some air out. When the compression is removed, some of the liquid rises into the chamber of the syringe. The small hydrometer floats in this liquid in a vertical position, its scale level indicating the specific gravity of the electrolyte. The hydrometer scale is graduated from 1.10 at the top to 1.30 at the bottom. A reading of 1.29 indicates a full charge.

EXERCISES. **37.** The density of an acid solution selected for a laboratory experiment was tested by the specific gravity bottle method. Find the specific gravity and density of the solution from the following data: Weight of the bottle when dry and empty, 15 g; weight of bottle filled with acid solution, 79 g; bottle filled with water, 65 g.

38. A milk bottle weighed on a trip-beam laboratory balance was found to have a weight of 771.2 g. When the bottle was filled with water, its weight was 1678.8 g, and when filled with milk, 1705.3 g. What was the specific gravity of the milk, as determined by this method?

39. A 100-gram balance weight is found to weigh 88.2 g in water and 86.5 g in a salt solution. What is the density of the solution?

40. A rectangular wooden rod 60 cm in length and having a cross section area of 1 cm^2, is weighted at one end so that it floats upright in water. The weight of the rod in air is 42 g. It is placed in water contained in a tall jar. (a) How many grams of water does the rod displace? (b) How many cm of the rod are below the surface? (c) If the rod is placed in gasoline, to what depth will it sink?

41. The rod mentioned in exercise 40 is placed in a salt solution. It sinks to a depth of 35 cm. What is the specific gravity of the solution?

SUMMARY

1. Mechanics of fluids deals with the effect of forces which are exerted on or by liquids and gases. Hydrostatics treats of the pressure of liquids at rest. Hydraulics treats of liquids in motion. Pneumatics treats of the pressure and motions of gases.

2. Density is mass per unit volume, that is, $D = M/V$.

3. Pressure is force per unit area, $P = F/A$.

4. Pressure in liquids (a) is proportional to depth, (b) is proportional to the density of the liquid.

5. Pressure in liquids at a given depth is equal to the height of the liquid column times the density, $P = HD$.

6. The force exerted by a liquid on the bottom of a vessel is equal to the area of the bottom times the pressure. $F = AHD$, where H is depth.

7. The force exerted by a liquid on the vertical side of a vessel is equal to the area of the side $\times$ ½ depth $\times$ density.

8. The force exerted by a liquid upon any submerged surface is equal to the area $\times$ the vertical distance of the surface from the center of area of the figure $\times$ the density, $F = AHD$, where H in this case is the vertical distance from the surface of the liquid to the center of figure of the body pressed upon.

9. A liquid in a system of communicating tubes rises to a common level in all the tubes.

10. Pascal's law: Pressure applied to a fluid confined in a closed vessel is transmitted equally in all directions to all parts of the fluid.

11. Archimedes' principle: A body immersed in a fluid is buoyed up by a force equal to the weight of the fluid displaced.

12. Archimedes' principle applies to bodies that float as well as to bodies that sink. A body that sinks displaces its own volume; a body that floats displaces its own weight. In both cases the body is buoyed up by a force equal to the weight of the fluid displaced.

13. Density is mass per unit volume; that is, $D = M/V$. When we speak of the density of a substance, we always say so many grams per cubic centimeter or so many pounds per cubic foot.

14. Specific gravity is the ratio of the density of a body to that of some substance taken as a standard. The standard chosen for solids and liquids is the weight of unit volume (cu. ft. or cm^3) of pure water. The standard for gases is air or oxygen.

15. When expressed in metric units, specific gravity and density are numerically equal to each other.

REVIEW EXERCISES

A

1. (a) What is the distinguishing property of a fluid as compared with that of a solid? (b) What is the distinction between a liquid and a gas? (c) What is the distinction between a vapor and a fog?

2. What is the density of water in (a) g per cm^3? (b) lb. per cu. ft?

3. In what units is density usually expressed in Density Tables?

4. What is the density of mercury? A cubical vessel 1 ft. on each edge is filled with mercury. What is the mass of mercury in pounds?

5. (a) Explain the meaning of each term in the equation $D = M/V$. (b) In what units is V expressed when M is given in grams? in pounds?

6. A 100-gram brass weight belonging to a beam balance set has what volume in cubic centimeters, the density of brass being 8.5 g/cm^3?

7. (a) What is the mass of a cu. ft. of brass? (b) What is the volume of a brass casting whose mass is 795.6 lb.?

8. What is the distinction between force and pressure?

9. Consider the vessel described in Exercise 4. What is the pressure exerted on the bottom of the vessel by the mercury in (a) lb./sq. ft. (b) lb./sq. in.?

10. Properties of fluids. (a) What is the direction in which a fluid exerts pressure at any point on the surface of the containing vessel? (b) What of the pressure at any point within the fluid? (c) What of the elasticity of fluids? (d) Compare the compressibility of gases and liquids.

11. Upon what does the pressure at any point in a liquid depend?

12. Explain each term in the equation $P = HD$, giving special attention to the meaning of H.

13. What is the pressure in gasoline (a) 10 cm below the surface? (b) 10 ft. below the surface?

14. A cubical vessel 10 cm on each edge is filled with water and placed with its bottom in a horizontal position. What force is exerted by the liquid (a) on the bottom? (b) on one side?

15. A cubical vessel is 10 ft. on each edge and is filled with water. What is the force (a) on the bottom, (b) on one side?

16. Under what conditions is the force on the bottom equal to (a) the weight of a liquid? (b) greater than the weight of the liquid? (c) less?

17. State Pascal's law, and give an illustration of its application.

18. Make a sketch to illustrate the principle of the hydrostatic press, and explain.

19. When force is applied to the small piston of a hydrostatic press, what do we get at the large piston, a multiplication of force or speed?

20. When the large piston of a hydrostatic press moves downward 1/10 in., the small piston moves upward 1 ft. A force of 50 lb. applied to the small piston will exert what force on the large one?

21. State Archimedes' principle, and explain how it may be illustrated.

22. It is desired to find the volume of a glass stopper of a bottle. Its weight in air is 9.1 g and in water 5.6 g. What is its volume?

23. A given body has a volume of 1 cu. ft. and a weight of 60 lb. Will it sink or float when thrown into (a) water? (b) gasoline? Explain.

24. Explain fully the difference between specific gravity and density.

25. By referring to Density Table VII, Appendix, how can you determine the specific gravity of any substance there listed? For example what is the specific gravity of flint glass?

B

26. Explain each term in the equation $F = AHD$, giving special attention to the meaning of H.

27. What is H in terms of the depth of the liquid when we are finding force on (a) the bottom of a rectangular vessel? (b) on one side?

28. By means of a sketch explain how an artesian well works.

29. Why is the Pelton water wheel more efficient than a water wheel having flat blades?

30. Explain briefly the principle upon which a water turbine works.

31. It is desired to design a hydrostatic press that will exert a force of 20 tons on the large piston when a pressure of 50 lb./sq.

in. is applied to the small piston. What must be the ratio of the cross sectional areas of the two pistons?

32. A small brass ball has a volume of 8 cm^3. What is its weight (a) in air? (b) in water? (c) in glycerine?

33. The dimensions of a rectangular metal block are 8 x 5 x 4 cm. Its weight in air is 1360 g; and in water, 1200 g. Find its density by two methods, that is, by the relation of mass to volume and the specific gravity method. How do the results obtained by these two methods compare? Which in this case is the shorter method?

C

34. A rectangular block of wood of density 0.8 g/cm^3 is 30 cm long, 10 cm wide, and 5 cm thick. When placed in water, it floats with its larger face parallel to, but above the surface. (a) What is its weight in air? (b) How much water does it displace? (c) What fraction of the block is above the water? (d) What is the relation between the fraction of the block submerged to the ratio of the density of the wood to that of water?

35. The specific gravity of ice is 0.9 and that of sea water, 1.03. What fraction of an iceberg floats out of water?

36. The specific gravity of a given liquid was determined by the specific gravity bottle method. The weight of the bottle was 15 g. The bottle was filled with water and weighed, after which it was filled with a given liquid. Weight of bottle filled with liquid, 78 g. The specific gravity was found to be 1.26. (a) What was the weight of the water? (b) What was the capacity of the bottle?

37. Describe the hydrometer, and give some of its uses. Why does the scale of a battery hydrometer read from the top downwards?

For additional problems, see Appendix.

Stop for test.

CHAPTER III

MECHANICS OF FLUIDS (Continued)

Atmospheric Pressure

63. The Atmosphere. The air of which our atmosphere is composed consists mainly of two gases, oxygen (O) and nitrogen (N) in very nearly the ratio by volume of one part of oxygen to four parts of nitrogen. In addition to oxygen and nitrogen, air also contains small quantities of other gases and varying amounts of water vapor. The most important physical properties of the air are as follows: (a) Air like all fluids is perfectly elastic; and (b) like all gases it is highly compressible, as is illustrated daily in the compression of air in the "pumping up" of bicycle and automobile tires. Liquids, we have learned, are practically incompressible, and hence pressure in them varies as the depth. Air, on the other hand, is highly compressible, and as a consequence most of the atmosphere lies very near the surface of the earth. At a height of three miles (altitude of Mt. Blanc) the density of the air is only about one-half as great as at sea level, and at an altitude of seven miles it is only one-fourth as great. This means that if a person were to go up to a height of seven miles, three-fourths of the atmosphere would lie below him.

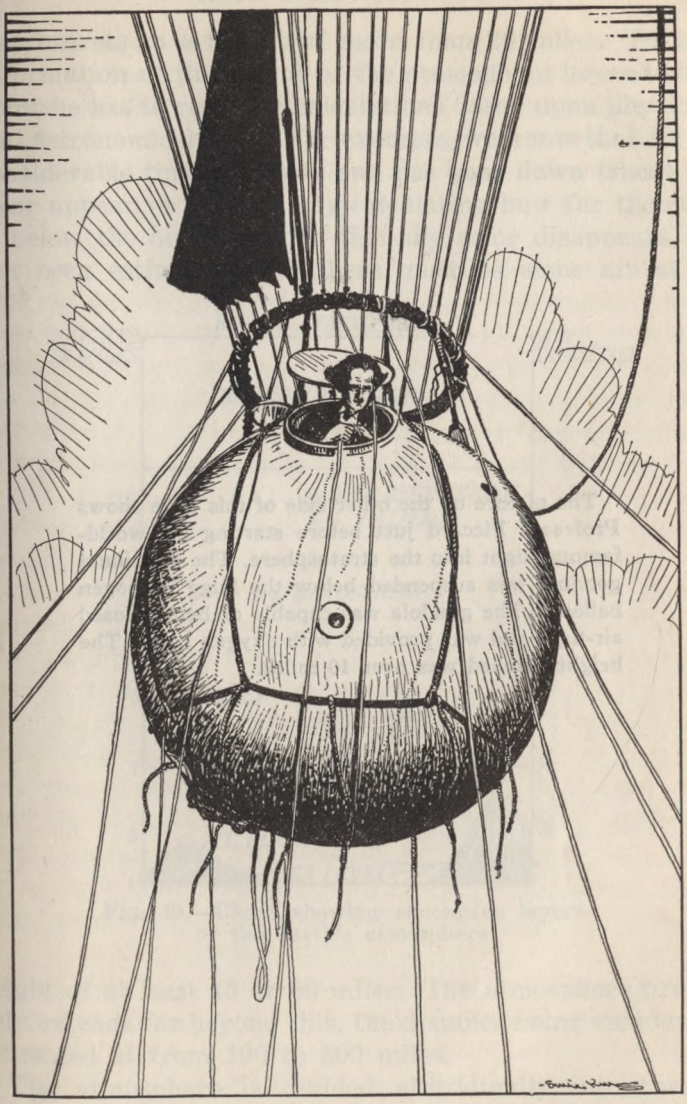
Children frequently ask the question, "What's at the top of the sky?" This is precisely the question that is today engaging the attention of scientists the world over. The spirit of science and adventure has led man to explore every region of the earth's surface. He is now turning his attention to the sky, that is, to the upper

regions of the atmosphere. Here is a world of silence and intense cold. No matter what storms may be raging upon the earth below, in these upper regions there is perfect calm. It is also a world devoid of life, except for the occasional adventurous visits of men. The present altitude record for any living thing, except man, is thought to be that of a certain species of Asiatic crow that explorers have found flying near the summit of Mt. Everest at a height of more than 25,000 feet. Investigation into the nature of this upper air world has thus far been made in three ways, namely, (a) by the use of airplanes and balloons controlled directly by man, (b) by the use of altitude balloons which are provided with self-regulating devices, and (c) by calculations based upon physical and astronomical data.

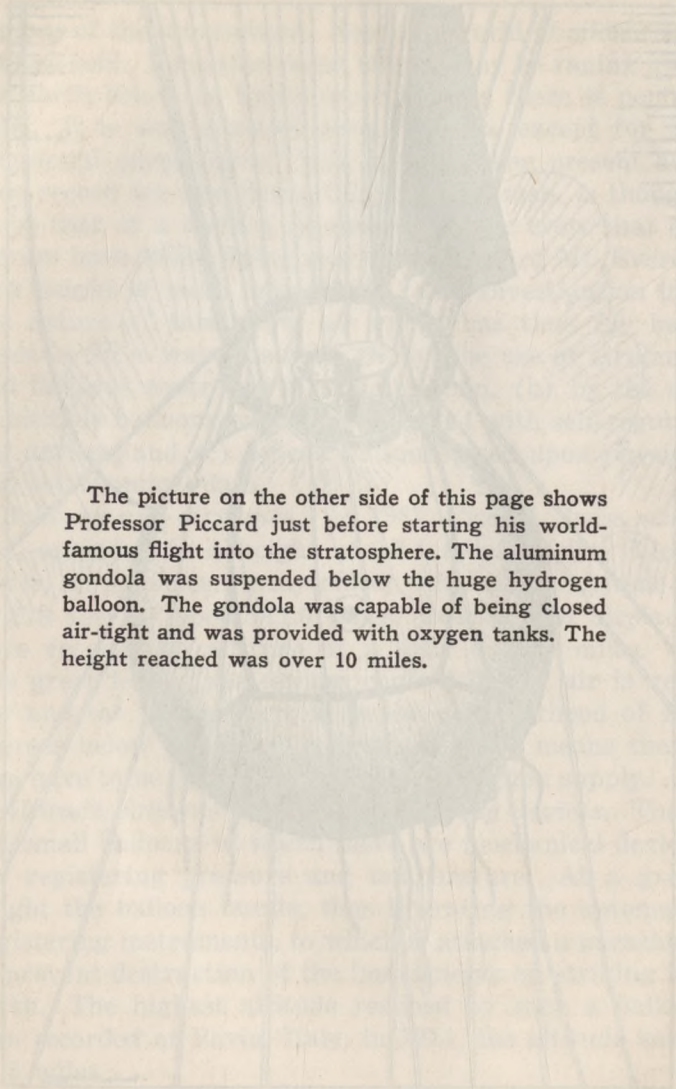
Balloons and airplanes. In 1931, Piccard, Swiss scientist, ascended to a height of 10.4 miles. Later Lieut. Settle, an American army officer, reached an altitude of 61,273 feet or about 11.6 miles. Man-piloted airplanes have reached an altitude of a little over 9 miles. At this great height the oxygen content of the air is very low and the temperature is in the neighborhood of 100 degrees below the freezing point. Special means therefore have to be provided for heat and oxygen supply.

Altitude balloons with self registering devices. These are small balloons in which there are mechanical devices for registering pressure and temperature. At a great height the balloon bursts, thus liberating the automatic registering instruments, to which is attached a parachute to prevent destruction of the instruments on striking the earth. The highest altitude reached by such a balloon was recorded at Pavia, Italy, in 1911, the altitude being 21.8 miles.

Calculated records. Man has explored the atmosphere either by direct observation or by means of self-recording



READY FOR A 10-MILE TRIP INTO THE STRATOSPHERE



The picture on the other side of this page shows Professor Piccard just before starting his world-famous flight into the stratosphere. The aluminum gondola was suspended below the huge hydrogen balloon. The gondola was capable of being closed air-tight and was provided with oxygen tanks. The height reached was over 10 miles.

instruments to a height of more than 20 miles. To get information of the extent of the atmosphere beyond this point he has to resort to calculations based upon physical and astronomical data. For example, we know that for a considerable time after the sun has gone down traces of color appear in the sky. By calculating how far the sun is below the horizon when this sky color disappears, it has been estimated that there must be some air at a

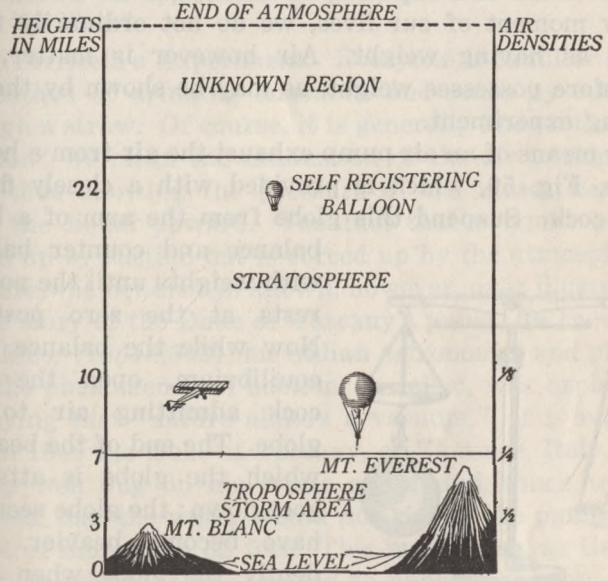


FIG. 49.—Chart showing successive layers of the earth's atmosphere

height of at least 45 or 50 miles. The atmosphere probably extends far beyond this, the distance being variously estimated at from 100 to 500 miles.

The atmosphere is divided altitudinally into three major zones, Fig. 49. The first is called the *troposphere*, a layer extending from the surface of the earth to a

height of some seven miles. We live at the bottom of this layer. This is the region of clouds and winds and in this zone there is packed three-fourths of the atmosphere. The second layer is the *stratosphere,* the upper limits of which are from 40 to 50 miles above the earth's surface. The third layer includes the *unknown region* which extends out to the limits of the atmosphere.

64. Air Has Weight. Since we live at the bottom of an ocean of air and experience its buoyant effects upon us every moment of our lives, we do not ordinarily think of it as having weight. Air however is matter, and therefore possesses weight, as may be shown by the following experiment.

By means of an air pump exhaust the air from a hollow globe, Fig. 50, which is provided with a closely fitting stop cock. Suspend this globe from the arm of a beam balance and counter balance with weights until the pointer rests at the zero position. Now while the balance is in equilibrium, open the stop cock admitting air to the globe. The end of the beam to which the globe is attached goes down; the globe seems to have become heavier. Evidently the globe when filled with air weighs more than when empty.

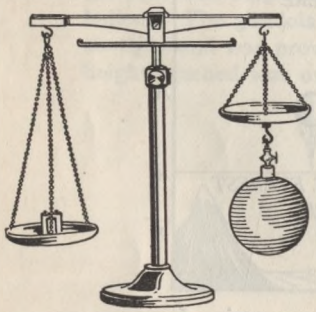


FIG. 50.—Weighing the air

The mass, and therefore the weight, of a unit volume of air depends upon the density of the atmosphere, which varies from day to day and from point to point on the earth's surface. Under standard conditions, that is at sea level and zero degrees Centigrade (0° C), the freezing point of water, the mass per unit volume is

1 *cubic centimeter of air* = 0.00129 *gram*
1 *cubic foot of air* = 0.08 *pound*

EXERCISES. **1.** The dimensions of an auditorium are 120 x 80 x 20 ft. When the density of air is 0.08 lb. per cu. ft., what is the weight of the air in this room, in tons?

2. A 1000 cc flask is weighed when full of air under standard conditions (pressure = 1 atmosphere at 0° C), and is weighed again after the air has been exhausted. Find the loss in weight, in grams.

3. What is the approximate weight of a cu. ft. of air at the top of Mt. Blanc?

65. Torricelli's Experiment. Everyone is familiar with the method of drinking lemonade and sodas by suction through a straw. Of course, it is generally known that by "suction" we mean the exhaustion of the air within the tube, thus allowing the pressure of the atmosphere to force the liquid upward. The fact that a liquid is not drawn up by suction but is forced up by the atmospheric pressure was not always known, however, as is illustrated by the story of the Duke of Tuscany's pump. In the days of Galileo, (1564-1642), an Italian astronomer and physicist, the phenomenon of suction, so called, was explained by saying that "nature abhors a vacuum." It is related that in 1640 the Duke of Tuscany, of Florence, Italy, had a deep well dug on his estate and found, much to his surprise, that the water would not rise in the pump to a height of more than 32 feet. This was a case, as Galileo put it, in which nature seemed to abhor a vacuum only to the height of about 32 feet. Torricelli, a young Italian scientist and a pupil of Galileo, undertook to solve the problem of the mysterious behavior of the water in the Duke's deep well pump. He believed that the reason the water did not rise to a height greater than 32 feet was because the pressure of the atmosphere was not great enough to support a column of water beyond this point. To test his theory he performed an experiment which

was destined to become a classic. Torricelli reasoned that if water will rise in a tube to a height of 32 ft., then mercury, which is 13.6 times as heavy as water, should rise only about 1/13 as high, that is, to a height of about 30 in.

Torricelli's famous experiment may be performed in the following manner: A glass tube about 80 centimeters in length, and sealed at one end, is filled with mercury. The open end is closed with the finger and the tube inverted, as shown in Fig. 51. The temporarily closed end is then placed under mercury in the dish and the finger removed. The mercury in the tube falls until the pressure due to the column of mercury is just counterbalanced by the pressure of the air upon the surface of the mercury in the dish. The space above the mercury in the tube is called a Torricellian vacuum. It contains a small amount of mercury vapor. At sea level the height of the mercury column is about 30 in., or 76 centimeters.

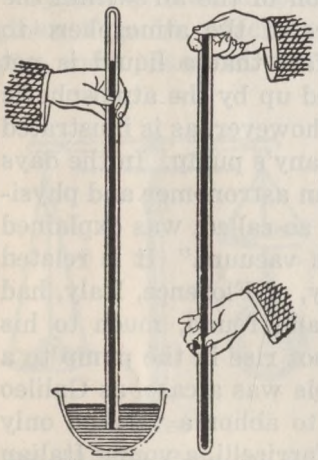


FIG. 51.—Torricelli's experiment

Torricelli concluded that the rise of liquids in tubes from which the air has been exhausted is due to the pressure exerted by the atmosphere on the free surface of the liquid and not due to any suction or pulling force resident in the vacuum. His experiment demonstrated in a satisfactory manner also that the failure of the water to rise to the surface in a deep well is not due to the presence of evil spirits as was thought, but is due to the fact that the pressure of the atmosphere is sufficient to support a col-

umn of water not greater than 34 feet in height at sea level.

66. Air Pressure Experiments. (a) *The tumbler experiment.* Atmospheric pressure may be strikingly illustrated by the tumbler experiment. In this case a tumbler is filled with water and a sheet of paper placed flat-wise upon the surface. Now place the palm of one hand upon the paper and with the other hand invert the tumbler. Remove the hand supporting the paper and water. Both remain in position, as shown in Fig. 52.



FIG. 52.—Air pressure keeps water in tumbler

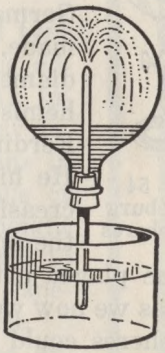


FIG. 53.—Fountain in vacuo

(b) *The vacuum fountain.* Into the neck of a flask, Fig. 53, fit a rubber stopper with one hole. Pass through this a short glass tube with the inner end drawn down to a fine opening. Fit one end of a rubber tube to the outer end of the glass tube and connect the other end of the rubber tube to an air pump. Exhaust the air from the flask. Pinch the rubber tube together, and insert the free end below the surface of the water in the dish. The pressure of the air upon the surface of the water will force a stream through the tube, forming a fountain inside the flask.

(c) *Magdeburg hemispheres.* One of the most widely known and most frequently mentioned experiments in physics is that with the Magdeburg hemispheres. These two hollow hemispheres with air-tight joint, Fig. 54, are pressed together and the air exhausted by means of a pump. The force holding the hemispheres together is equal to a pressure of 14.7 pounds per square inch multiplied by the cross sectional area through the center of the sphere.



FIG. 54
Magdeburg hemispheres

Otto von Güricke, mayor of Magdeburg, Germany, and inventor of the hemispheres, made, in 1654, a pair of them 22 inches in diameter. At Ratesbon he exhibited the hemispheres before the Emperor of Austria, Ferdinand III, and the imperial parliament. He hitched horses to each hemisphere, increasing the total number to sixteen before the spheres were pulled apart. In this respect he played a trick upon his wondering guests, because, as we now well know, and as he no doubt knew, the hemispheres could have been pulled apart by fastening one side of the hitching apparatus to a stake or a tree and the horses to the other side, in which case only eight horses would have been required. The experiment, however, would not have been so striking.

EXERCISES. **4.** Explain why the liquid in a medicine dropper will not run out until the bulb is squeezed.

5. So long as the flask fountain shown in Fig. 53 continues to play, what can you say of the air pressure inside the flask as compared with that outside? What about the air pressure inside and out when the fountain ceases to operate?

6. Let us assume that the atmospheric pressure was 14.7 lb. per sq. in. at the place where von Güricke performed his experiment. The force required to pull the hemispheres apart was equal to the weight of a column of air having a cross sectional area equal to that of the sphere formed by the two hemispheres when

fitted together, and equivalent to a pressure of 14.7 lb. per sq. in. What was the force required to pull the hemisphere apart?

67. The Mercury Barometer. The *mercury barometer* is a Torricellian tube fastened to a frame having a scale, by means of which the height of the mercury may be read. If such a barometer be carried up a mountain side, the mercurial column will fall, due to the diminished pressure of the atmosphere; if it be taken down into a deep mine or valley, the mercurial column will rise.

A mercury barometer of standard type is shown in Fig. 55. The scale V is adjusted by means of the screw S. An instrument of this sort furnishes a very accurate means of measuring variations in atmospheric pressure, since by means of the sliding scale V at the top, the height of the mercurial column may be read to the fraction of a millimeter. It is especially well suited for laboratory work.

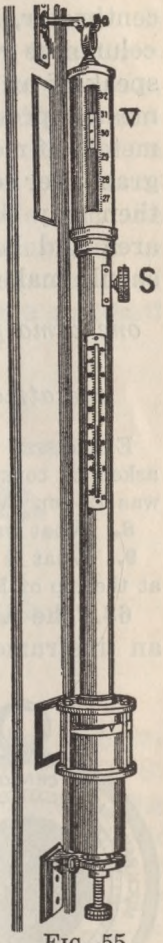


FIG. 55
Standard barometer

Pressure of one atmosphere. The expression "a pressure of one atmosphere" means the pressure exerted by the atmosphere at sea level and 0° C. Under standard conditions (sea level, 0° C) the height of the barometric column is 76 centimeters, which is equivalent to about 30 inches. A column of mercury 76 centimeters in height and having a cross sectional area of 1 square centimeter has a mass of 1033.3 grams. A pressure of one atmosphere, therefore, is equal to 1033.3 grams per square centimeter, which is equivalent to 14.7 pounds per square inch.

It is customary in laboratory practice to speak of atmospheric pressure not as so many grams per square centimeter, but in terms of the height of the mercury column as registered by the barometer. Thus, when we speak of an atmospheric pressure of 74 centimeters, we mean a pressure that will support a column of 74 centimeters of mercury, which is equivalent to 74/76 of 1033.3 grams per square centimeter. Since atmospheric pressure then may be expressed both in terms of force per unit area and height of the mercurial column, the student should make note of the following:

one atmosphere, pressure equals { 1033.3 *g/cm²* / 14.7 *lb./sq. in.*

one atmosphere, equivalent to { 76 *cm mercury* / 30 *in. mercury*

EXERCISES. **7.** On a certain day a laboratory attendant was asked to consult the barometer. He reported that the pressure was 72 cm. What was the atmospheric pressure in g/cm²?

8. What was the pressure (Ex. 7) in lb./sq. in.?

9. What is the approximate barometric reading in centimeters at the top of Mt. Blanc? See Art. 63.

68. The Aneroid Barometer. An *aneroid barometer* is an instrument which indicates atmospheric pressure changes by the movement of a pointer over a scale, Fig. 56. It contains no mercury or other liquid. The principle working part consists of a small air tight cylindrical metallic box with a thin cover which is somewhat curved. Change of pressure causes the curvature of the top to change slightly. This motion of the cover up or down as the pressure decreases or increases is communicated to the pointer



FIG. 56
Aneroid barometer

by means of a system of delicately adjusted levers.

The scale readings of an aneroid barometer are determined by comparing it with a mercury barometer. It is not so accurate as the mercury barometer; it is, however, very much more convenient for certain purposes, since because of its shape and size it may readily be carried about from place to place. It is very sensitive; a good aneroid barometer registers differences of atmospheric pressure between the top of a table and the floor. In order to avoid inaccuracy in reading, due to disarrangement of the lever system, an instrument of this sort has to be checked up frequently with a standard mercury barometer.

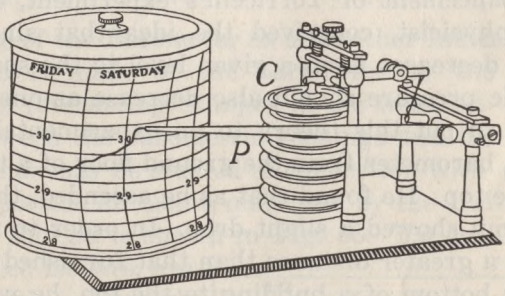


FIG. 57.—Barograph

A self-recording aneroid barometer is called a *barograph*, Fig. 57. This instrument is used by the U. S. Weather Bureau in recording pressure changes. The paper record sheet is wrapped around a drum. Clock work inside the drum makes it revolve. The cylinder *C* is the air tight box, the changing shape of the cover of which moves the lever system. The recording pointer *P* is provided with a small fountain pen. As the pointer moves up or down due to pressure changes, the pen makes a record on the revolving paper record sheet. The clock work is so set that it takes a week for the drum to

make one complete revolution. At the end of the week, or whatever time may have been decided upon, the record sheet is removed. The curve traced upon it gives a continuous record of the barometric changes for the period.

69. Use of the Barometer in Determining Altitude. The barometer is today used quite extensively by geological and surveying parties to compute the altitude of different places above sea level. It is also used by aviators in determining elevation above the earth. The use of the barometer in making altitude determinations has back of it an interesting story. It will be recalled that Torricelli was one of the first to demonstrate experimentally that the atmosphere exerts pressure. Some five years after the announcement of Torricelli's experiment, Pascal, a French physicist, conceived the idea that since water pressure decreases from a given level to the surface, atmospheric pressure should also decrease as one ascends. In order to put this theory to an experimental test, he carried a barometer from the ground floor of a tall building to the top. He found that as he ascended, the mercurial column showed a slight drop. In order to make the test over a greater distance than that furnished by going from the bottom of a building to the top, he wrote to a friend who lived in southern France near the mountain Puy-de-Dome (pronounced *Pwe-de-dome*), asking him to carry a barometer up the mountain side and to note the fall of the mercury as he ascended. The experiment was performed according to Pascal's instructions, the friend ascending the mountain to a vertical height of about 1000 meters. He observed with "admiration and astonishment" that the column had fallen a distance of 8 centimeters. Thus it was demonstrated that a decrease of the barometric reading corresponds to an increase in altitude.

At places near the sea level a fall of 0.1 centimeter in the barometric column corresponds to a vertical ascent

of about 12 meters; a fall of 0.1 inch, to about 90 feet. This rate of decrease is not maintained, however, on reaching high levels, owing to the rapidly decreasing density of the air. Accordingly, altitude tables have been prepared which make allowance for the lessening air pressure and also for temperature changes.

EXERCISES. **10.** It is desired to measure, approximately, the height of a hill by means of an aneroid barometer. The barometric reading at the bottom of the hill is 75.4 cm; at the top, 74.8 cm. What is the height of the hill, in ft.?

11. Measure the distance from the basement to the topmost floor of your school building. This distance is equivalent to what fall in barometric reading? If an aneroid barometer is available, compare your computed value with the actual fall as determined by carrying the instrument from the basement to the upper floor.

70. Use of the Barometer as a Weather Indicator. Constant use is made of the barometer by the Weather Bureau in forecasting changes in the weather. The relations between the barometric readings and the direction of the winds may be studied in connection with a weather map. A portion of such a map used by the United States Weather Bureau is shown in Fig. 58. The heavy curved lines, called isobars, are lines passing through places of equal atmospheric pressure. At the place marked *LOW* the atmospheric pressure is least, while at the place marked *HIGH* the pressure is greatest. The air flows into this "low" region, forming a sort of whirlpool. In the northern hemisphere the general direction of the wind about these areas of low pres-

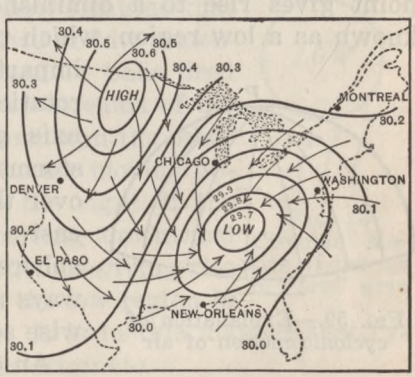


FIG. 58.—Weather map

sure is counter-clockwise, that is, in the opposite direction to motion of the hands of a clock.

The observations of the Weather Bureau on barometric pressure for a series of years indicate that these low pressure areas are continually passing over the country with considerable regularity and along pretty well defined paths. The forecaster depends largely upon his knowledge of the motion of these areas in predicting the weather.

The relation of the barometric reading to the probable condition of the weather may be stated as follows: (a) A rising barometer precedes fair weather; (b) a falling barometer precedes foul weather; (c) a sudden fall in the barometer indicates a storm; (d) a steady barometer indicates settled fair weather.

71. Circulatory Motion of the Air. Few fluid motions are of greater importance to us than those of the air. Especially is this true of what is known to the meteorologist as "cyclones," those immense whirls of air many miles in diameter, which give rise to the varying conditions of the "weather." The air being heated at a given point gives rise to a diminished barometric pressure, known as a low region, which together with the motion imparted to the air due to the rotation of the earth on its axis, gives rise to the cyclonic storms which pass periodically over the country from west to east. In the northern hemisphere these wind storms always rotate in a counter-clockwise sense.

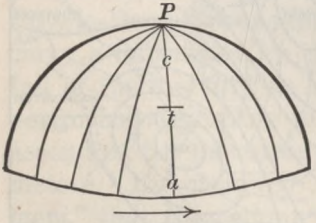


FIG. 59.—Explanation of cyclonic motion of air

An explanation of this counter-clockwise rotation may be given as follows: Consider the three points *a*, *t*, *c*, on a meridian of the hemi-

sphere, Fig. 59, which rotates about a vertical axis through P in the sense indicated by the arrow, that is, from west to east. Suppose that t is a target, pivoted so as to rotate about its center, and that guns situated at a and c are aimed directly at its center. Considering the motion of the sphere, we wish to find where the balls fired from the guns will strike the target, with reference to its center, all three points moving in the same direction but with different speeds.

Now if the hemisphere were at rest, the balls fired from a and c would strike the target t at its center, but on opposite sides. But the hemisphere is not at rest; it is rotating from west to east. It is evident then that the speed of a is greater than that of t, and hence the ball from a will strike t to the right of the center of the target. Likewise, the speed of c is less than that of t, and, therefore, the ball from c will strike t to the left of the center, assuming that we are looking from a toward the pole P. This means that the target t will tend to rotate about its center in a counter-clockwise sense, due to the impact of the two balls. Let us now imagine that the region about t is a place of low atmospheric pressure, and instead of balls fired from a and c, we have two streams of air rushing in, one moving north and the other moving south, Fig. 60. The air moving upward will be deflected to the right of the low area, and the air moving downward will be deflected to the left, thus giving rise to a circulatory motion in the low region in a counter-clockwise sense. It is in this manner that cyclonic rotations of the air originate.

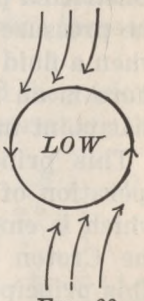


FIG. 60
Circular motion of the air in "low" area

72. Velocity and Pressure in Fluids. When a stream of air or water, or any other fluid, encounters a region

of constricted cross sectional area, as at p, Fig. 61, there occurs in this region an *increase* in the velocity of the fluid and a *decrease* in lateral pressure. That this rather surprising condition with reference to increase in velocity and decrease in pressure is true may be demonstrated by allowing a stream of water under constant pressure to flow through a horizontal tube to which there are attached three vertical tubes, p', p, p'', Fig. 61. The lateral pressure at the respective points exerted by the stream of water flowing horizontally through the pipe is measured by the height of the liquid in each of the vertical tubes. The velocities in the enlarged portions of the tube are the same, that is, the pressure at p' equals that of p''. In the constricted portion there is an increase in velocity. Here the pressure p is less than at p' or p''. This means that when a fluid in motion is constricted at any point so that there is an increase in velocity, the lateral pressure at that point in the fluid is diminished.

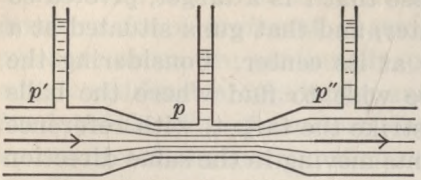


FIG. 61.—The pressure is least where the velocity is greatest

This principle has an important application in the operation of the Venturi water meter, a modification of which is employed in measuring water passing through the Croton Aqueduct which supplies New York City. This principle may also be employed to explain the operation of the jet pump, the atomizer or spraying machine, the card and tube experiment, the curving of a pitched ball, and certain phases of the operation of the airplane, as will be explained later.

73. Card and Tube Experiment. If air be blown through a tube having a flaring lip ab, and a card C be placed against the lower opening, as shown in Fig. 62, it

will be found that the card cannot be blown from the tube, even if the tube be held in a vertical position. This phenomenon may be explained in terms of the principle of diminished pressure as illustrated in the case of the flow of water in a constricted pipe, Fig. 61. The air passing downward through the tube issues with high velocity through the narrow opening between the card and the flaring lip *ab*. As the velocity of the air increases its lateral pressure diminishes. The pressure of the atmosphere upon the bottom, therefore, holds it in place.

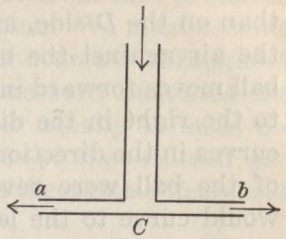


FIG. 62.—Card-and-tube experiment

74. The Curving of a Pitched Ball. To every boy interested in baseball there has occurred the question, Why does a pitched ball curve? The answer to this question involves the principle just discussed in Art. 72, namely, the principle of diminished lateral pressure with increase in velocity.

Consider Fig. 63. Here we have illustrated the various motions of a pitched ball. The line *A R B* represents the general direction in which the ball is moving at a given instant after leaving the hand of the pitcher. Two motions are imparted to it by the pitcher, one in the direction *A B*, that is toward the batter, and the other a rotary motion *R*, in a clockwise sense.

We may consider the ball as moving against the air or, on the other hand, we may consider the air as moving against the ball with a speed represented by the lines *D* and *E*. As the

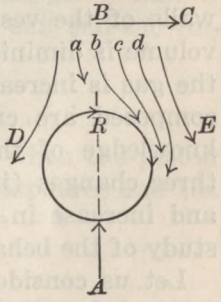


FIG. 63.—Curving of a pitched ball

ball rotates it carries air around with it as shown by the lines *abcd*. The velocity of the air on the *E*-side is greater than on the *D*-side, and therefore the lateral pressure of the air against the ball is less at *E* than at *D*. As the ball moves forward in the direction from *A* to *B*, it curves to the right in the direction *BC;* that is to say, the ball curves in the direction in which it rotates. If the rotation of the ball were reversed, that is counter-clockwise, it would curve to the left.

BOYLE'S LAW

75. Pressure, Volume, Density. We are now going to study the relation which exists between the pressure of a confined gas and its volume and the resulting density changes which occur when the pressure is varied. By way of illustration, let us consider that we have 1000 cubic centimeters of a gas (air say) confined in a cylindrical vessel which is provided with a closely fitting piston, Fig. 64. Suppose that force is exerted upon the piston sufficient to push it down from *A* to *B*, that is half way down the cylinder. Three things happen: (a) First, the pressure which the gas exerts upon the containing walls of the vessel is increased (doubled); second, the volume is diminished (halved); and third, the density of the gas is increased because the molecules of which it is composed are crowded more closely together. Now a knowledge of the relations which exist between these three changes (increase of pressure, decrease in volume, and increase in density) is of great importance in the study of the behavior of gases.

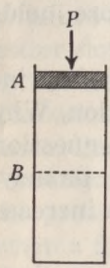


FIG. 64

Let us consider all the factors which enter into this suggested experiment. First, the experiment must be carried out under conditions of *constant temperature*, for

if the gas is cooled, for example, during the process, a contraction would result quite apart from that produced by the change in position of the piston. We must assume then that the temperature remains constant throughout the experiment. Second, as the pressure increases the volume decreases. And third, since density is mass divided by volume and since the mass of the enclosed gas remains constant it follows that as the *volume decreases* the *density increases.*

76. Boyle's Law. In 1660 Robert Boyle, famous Irish chemist and physicist, first investigated the exact relation between the volume of a gas and the pressure to which it is subjected. His results are stated in what we call *Boyle's law,* which may be stated thus: *The temperature remaining constant, the volume of a gas varies inversely as the pressure.*

Since the volume decreases in the same ratio as the pressure increases, this law is sometimes written in the form of an equation,

$$\textit{pressure} \times \textit{volume} = \textit{a constant}$$

$$PV = C$$

In order to explain the meaning of this equation, let us again consider the case illustrated by Fig. 64. When the piston is at A, the pressure on the enclosed gas is that of 1 atmosphere, that is, 1033.3 grams per cm^2 and the volume is 1000 cm^3. Then

$$PV = C$$

$$1 \times 1000 = 1000$$

in which the constant C is 1000. Now when the piston moves down to B, the pressure is 2 atmospheres and the volume becomes 500 cm^3. Then we have

$$P'V' = C$$

$$2 \times 500 = 1000$$

As before the value of C for this experiment remains equal to 1000.

Now since PV and $P'V'$ are equal to the same thing (C) they are therefore equal to each other, and hence

$$PV = P'V'$$

We have in this equation four factors, and therefore when any three are given, we can find the fourth.

Example. A gas enclosed in a cylindrical vessel provided with a piston has a volume of 5 cu. ft. when the pressure is 1 atmosphere, that is, 14.7 lb. per sq. in. What will be the volume of the gas when the piston moves down so as to increase the pressure to 20 lb. per sq. in.? *Solution*: Assuming that Boyle's law holds, we may write

$$PV = P'V'$$
$$14.7 \times 5 = 20 \times V'$$
$$\text{Hence } V' = \frac{14.7 \times 5}{20} = 3.675 \text{ cu. ft.}$$

77. Verification of Boyle's Law. Boyle's law and its equation may be verified by means of an apparatus similar to that shown in Fig. 65. Here we have two glass vessels, C and V, connected by means of a strong rubber tube. Both the vessels and the connecting tube contain mercury. To begin with, the mercury stands at the same level in both parts of the apparatus. The air in the chamber C, 20 cm³ say, is under a pressure of one atmosphere. Now applying the law we have

$$PV = 1 \times 20 = 20$$

Now suppose that the vessel V be elevated until the pressure upon the enclosed air is 2 atmospheres. The volume is now 10 cm³. Since we have a different pressure and a different volume to deal with, the equation in this case is written

$$P'V' = 2 \times 10 = 20$$

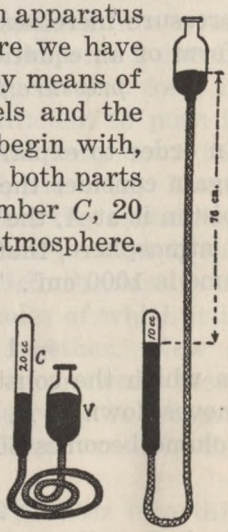


FIG. 65.—Boyle's law apparatus

If the pressure be increased to 4 atmospheres the volume will be decreased to 5 cm^3; hence

$$P''V'' = 4 \times 5 = 20$$

It will be noted that in every case the product of the pressure and the volume is the same; that is, PV equals C, in which the constant C in this particular case is 20.

EXERCISES. **12.** (a) A pressure of one atmosphere is equivalent to how many cm of mercury? (b) To what height must the mercury in vessel V, Fig. 65, be elevated to get a pressure of two atmospheres? (c) What height of mercury column above the level in C would be required to give a pressure of four atmospheres?

13. A pressure of four atmospheres is equivalent to how many grams per square centimeter?

14. A pressure of four atmospheres is equivalent to how many lb. per sq. in.?

15. A cylindrical metal tank is fitted with a piston, Fig. 64. When the piston is at the top of the tank, the volume of air enclosed is 2000 cm^3. Its pressure is one atmosphere, and its density 0.00129 g/cm^2. The piston is then pushed down until the air below it is compressed into a volume of 500 cm^3. (a) What is the pressure exerted by the compressed air, in g/cm^2? (b) What is the density?

16. A spherical vessel, provided with a stop cock and having a capacity of 1 cu. ft., is filled with air under a pressure of 10 atmospheres ($14.7 \times 10 = 147$ lb. per sq. in.) The stop cock is opened and the compressed air is allowed to escape until the pressure inside the vessel is equal to that of the air outside, namely, 14.7 lbs. per sq. in. What volume is occupied by the air which escaped?

78. Limitations of Boyle's Law. Boyle's law holds true for most gases under ordinary pressure and temperature conditions. Deviations from the law occur, however, when the pressure becomes very great, especially in the case of those gases which are easily liquefied, such as carbon dioxide (CO_2), sulphur dioxide (SO_2), and chlorine (Cl_2).

Within moderate limits of pressure, Boyle's law is exceedingly useful in the study of gases. It finds many

applications in physics and chemistry and in the solution of certain engineering problems.

79. Measurement of Gas Pressure. The measurement of gas and steam pressure is usually made by means of a pressure gauge. The Bourdon gauge is one of the most common and effective instruments of this type. A sectional drawing of the Bourdon gauge is shown in Fig. 66. The gauge is circular in form and is provided with a scale which is calibrated to give readings in pounds per square inch. Within the instrument there is a circular flattened tube *T*, one end, *G*, of which is designed to communicate with the gas the pressure of which is to be measured, and the other end is connected to a lever *L*, which operates the pointer *P*. When the gas, as for example steam from a boiler or air from an automobile tire, enters the circular tube *T*, it tends to straighten the tube, thus causing the pointer to move over the scale in a clockwise sense. The position at which the pointer comes to rest on the scale indicates the pressure in pounds per square inch.

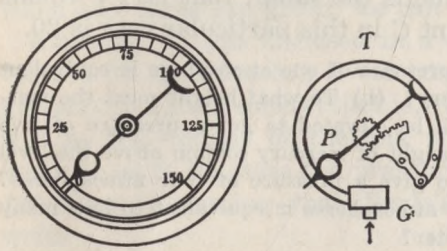


FIG. 66.—Bourdon pressure gauge



FIG. 67
Automobile tire pressure gauge

A spring gauge of the type shown in Fig. 67 is sometimes used by garage men to measure the air pressure in automobile tires. The gauge is pressed against the valve-end, making a tight joint and at the same time opening the valve allowing air to

enter at A. The air under pressure forces the spring, together with the cylindrical scale-tube, upward. When the gauge is detached from the tire, the spring recoils downward to its position of rest, leaving the scale-tube in its upper position. After the reading is taken, the scale-tube is pushed down to the zero position.

It should be noted that gauges of the sort described above record pressures in excess of atmospheric pressure. That is to say, the total or *absolute pressure* is the gauge reading plus 14.7 pounds per square inch at sea level.

EXERCISES. **17.** An automobile tire contains 1800 cu. in. of air under a gauge pressure of 35 lb. per sq. in. What is the absolute pressure exerted by the enclosed air upon the tire?

18. What volume would the air in the tire (Ex. 17) occupy at a pressure of one atmosphere?

19. Assign a member of the class to visit a garage and secure the following information: (a) the air pressure maintained in the charging tank; (b) air pressure supplied to different types of tires (light and heavy automobiles and trucks); (c) kind of pressure gauge used.

AIR PUMPS

80. The Exhaust Pump. Historically the air pump is one of the oldest pieces of physical apparatus which we have in our laboratories. We have learned (Art. 66c) something of von Güricke's famous experiment with the Magdeburg hemispheres. In order to exhaust the air from his hemispheres, it was necessary for him to design and construct an air pump, which he did in 1650. His pump was similar in principle to that of the simple two-valve exhaust pump used today. The principle upon which the valves of this air pump work is shown in the outline sketch of Fig. 68. Every pump of this type, whether for air or water, has two valves. On the downward stroke the piston valve V opens and the valve V' closes. On the upward stroke V closes and V' opens. This opening and closing of the valves may be explained

as follows: On the downward stroke of the piston containing valve V, the air in the chamber below it is somewhat compressed, causing valve V' to close. The compressed air is forced out through valve V as the piston moves downward. On the upward stroke of the piston valve V closes and the air in the receiver R expands, some of it flowing out through valve V'. Thus on each double stroke of the piston a certain fraction of the air is pumped out of the receiver R. This pumping process may be carried on until the pressure of the air in R is too low to open valve V'. The limit of exhaustion is now reached for a pump of this type.

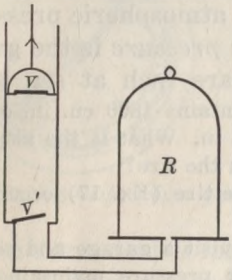


FIG. 68

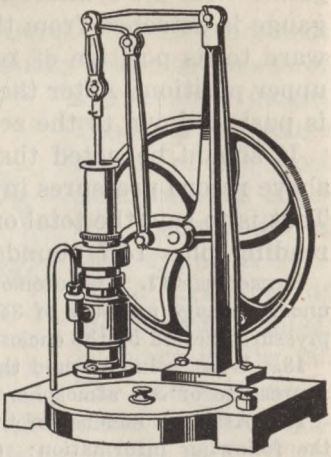


FIG. 69

Air pump

A modern type of laboratory air pump is shown in Fig. 69. This is a single cylinder pump the piston of which makes one exhausting stroke for each revolution in the fly wheel. This pump is provided with accurately ground metal valves which are immersed in oil, thus insuring a perfect fit.

EXERCISES. **20.** An air pump of the type shown in Fig. 69 is connected to a receiver which contains air under pressure of one atmosphere (76 cm of mercury). Suppose that on each complete revolution of the fly wheel the piston removes 1/10 of the air from

the receiver. After one revolution of the wheel (a) what fraction of the original quantity of the air remains in the receiver? (b) What is its pressure? (c) What is its density?

21. What is the pressure in g/cm² of the air in the receiver after 4 revolutions?

81. The Water Jet Pump. The *water jet pump* or *aspirator,* Fig. 70, is a form of pump adapted for use where water under high pressure is available, as from the city water mains. A stream of water under pressure enters the tube *a*, which is constricted at the lower end, thus giving the stream a high velocity at *b*. The air is drawn along with the stream of water in the form of small bubbles, thus exhausting the chamber *C*.

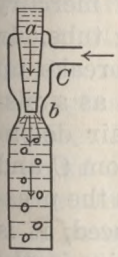


FIG. 70
Water jet pump

With a pump of this sort it is possible, with water from the city mains, to obtain a vacuum of about 5 cm of mercury. Since a pressure of one atmosphere is equivalent to a column of mercury 76 cm in height, a vacuum of 5 cm is equivalent to 5/76 of 1 atmosphere.

The water jet pump is very useful in laboratory and commercial operations in which only a partial vacuum is required. Water aspirators (jet pumps) are widely used by dentists in connection with mouth operations. The aspirator is hooked up to a rubber tube having a curved tube-like mouthpiece at one end. When the aspirator is in action and the curved glass tube is inserted into the patient's mouth, saliva is drawn out with the air, thus leaving the mouth comparatively dry.

82. The Mercury Air Pump. A good mechanical two-valve pump will exhaust the air from a vessel until the pressure within the receiver is equivalent to less than one millimeter. In order to get a better vacuum than this, it is necessary to use some form of the mercury air

pump. The principle upon which the mercury air pump works is exactly the same as that of the water jet pump, only mercury is used instead of water. These pumps are usually made of glass, mercury playing a part somewhat analogous to that of the piston in mechanical pumps. A sectional sketch of a mercury pump is shown in Fig. 71. A column of mercury is allowed to fall through a long tube. In passing through chamber *C* it breaks up into drops, each of which serves as a piston to carry a portion of the air downward, thus exhausting the air from *C* and from the attached receiver *R*. As the pressure within the receiver is reduced, it is obvious that the mercury will rise in the tube *B*, standing at a height of nearly 76 centimeters when the exhaustion has reached a maximum. Pumps of this type are often used in exhausting the air from electric light bulbs.



FIG. 71
A simple form of the mercury pump

The mercury air pump will produce a vacuum of less than one-millionth of an atmosphere.

83. The Mercury Vapor Air Pump. One of the latest and most effective air pumps is known as the Diffusion or Mercury Vapor Pump. This pump is more than a million times as effective as the ordinary two-valve laboratory pump. It consists of three pumps, *A*, *B*, *C*, connected in series. Pump *A* is a water aspirator and *B* and C are mercury pumps, in which the mercury is vaporized by means of electric furnaces. The three pumps are connected in series, that is, *A* is connected to *B* and *B* to *C*. Pump *C* draws the air from the receiver *R* through a tube containing liquid air, the temperature

of which is far below that of the freezing point of mercury.

The principle upon which the pump works is somewhat as follows: The first step in exhaustion is secured by means of the water jet pump *A*, which produces a vacuum of about 10 mm. Second, mercury in *B* is heated by means of an electric furnace. As the mercury vapor flows through the jet tube, the vacuum is reduced to about .01 mm. Third, mercury vapor in pump *C*, which is also heated by an electric furnace, flows toward *B* through a jet tube producing a very high vacuum in the receiver *R*. Air from *R* is drawn out and down through a liquid air chamber in order to freeze out any mercury vapor which might otherwise work up into *R*.

A pump of this sort will produce a vacuum of one-hundred-millionth of a millimeter. Diffusion mercury pumps are used in exhausting radio bulbs and other high vacuum apparatus.

84. The Ejector Principle. When a strong wind blows across the top of a chimney, Fig. 72, some of the air in the chimney is drawn out and a partial vacuum is created. As a result a strong upward draft is created in the chimney. The chimney, or for that matter any tube across the open end of which a stream of air is driven, illustrates what is called an *ejector principle.*

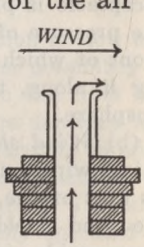


FIG. 72
Ejector principle in a chimney

If a strong current of air be blown across the opening of a tube, the air in the upper part of the tube will be dragged out by the moving column, and the pressure within thus reduced. If the lower end of the tube be immersed in a liquid to which the atmosphere has free access, the liquid will be forced to the mouth of the tube and blown into a fine spray. Such an instrument is

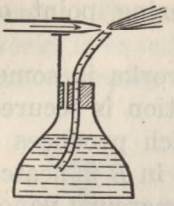


FIG. 73
Atomizer

called an *atomizer*, Fig. 73. It is used by physicians for spraying the throat.

The atomizer principle is commonly employed in spraying trees and shrubs. It is largely used also in spray pump painting. The spray pump when used for this purpose is sometimes called a "paint gun."

85. Application of the Vacuum Principle. A vacuum may be produced mechanically by means of a moving piston or by a rotating fan; it may also be produced by a column of fluid (water, mercury, or air) flowing through a chamber or across the open end of a tube. Practical application of the various methods are illustrated by the pneumatic tube, the automobile vacuum wind shield wiper, the vacuum milking machine, and the household vacuum cleaner.

(a) *Pneumatic tube systems.* In some department stores pneumatic dispatch tubes are used for carrying cash and mail to and from the central office. The air from these tubes is exhausted by a powerful air pump. The articles (cash or mail) to be sent are placed in small leather cases which fit the tube closely. When the receptacle is placed in the vacuum tube, it is driven through by the pressure of air behind it. We thus have a moving system in front of which there is a partial vacuum and behind which, pushing it along, there is a pressure equivalent to that of the atmosphere.

(b) *Wind shield wiper.* The operation of the pneumatic wind shield wiper, used on certain automobiles, is essentially the same as that of the pneumatic tube device, except that in the case of the wind shield wiper the pressure is applied first to one side of the movable piston and then to the other by means of a trip valve. The piston which operates the wiping lever is enclosed in a cylinder which is connected with the engine chamber by means of a tube. When the engine is running, a partial vacuum is produced. The piston is driven toward the vacuum end of the cylinder by air pressure. Let us suppose that the motion of the piston, just mentioned, is toward the left. Upon reaching the left end of its stroke, the piston automatically opens a valve which admits air to the left side of the cylinder and at the same time allows the engine to exhaust the air from the right side. Thus the piston

moves back and forth, the partial vacuum alternating from one side to the other.

(c) *Vacuum milking machine.* Until quite recently the milking

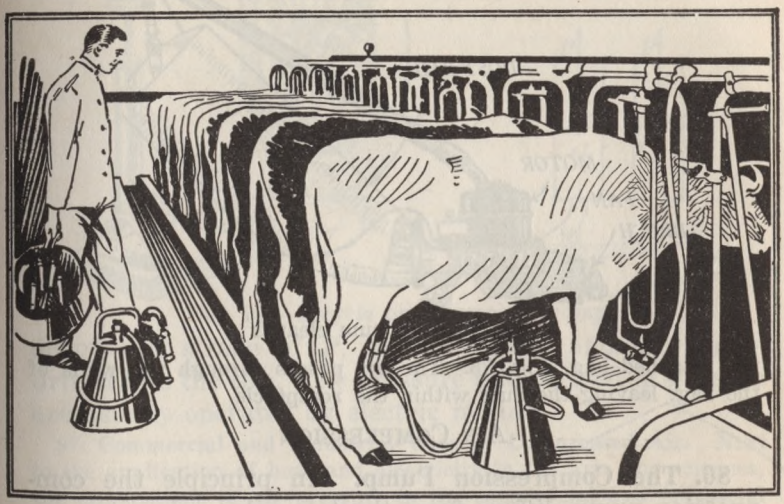


FIG. 74.—A modern dairy; milking by the vacuum process

of cows was everywhere done by hand. This operation is slow and laborious, and it is not infrequently attended by the possibility of contamination of the milk by dust and germs from the cows and from the hands and clothing of the milker. A more sanitary and efficient method is that which makes use of vacuum milking machines, Fig. 74, which in recent years have been installed in many of our first-class dairies. The essential feature of a milking machine is an evacuating air pump, which is usually operated by a small gasoline engine, and to which there is connected a long tube. One end of this tube is attached to a vessel for the reception of the milk. This vessel is provided with four tubes, the free ends of which are attached to the udder of the cow. The operation of the air pump evacuates the receiving vessel and the tube system, and the milk is drawn quickly and effectively from the cow.

(d) *Vacuum cleaner.* The vacuum cleaner is today almost as common an article in the household as the broom and dustpan of our grandmother's time. This apparatus is provided with a fan-like suction pump which is driven by a small electric motor. The

air, together with the dust and dirt from the carpet, is blown into

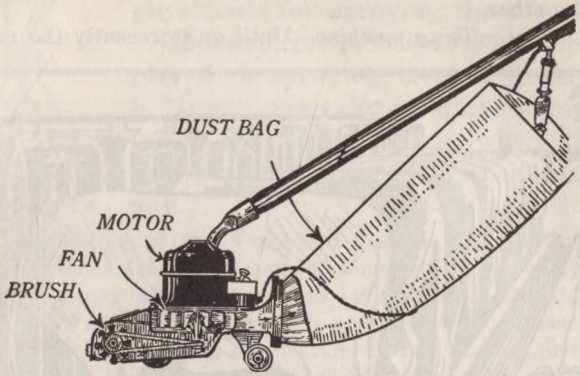


FIG. 75.—Vacuum cleaner

the attached bag, Fig. 75. The air passes through the walls of the bag, leaving the dust within the receptacle.

AIR COMPRESSION

86. The Compression Pump. In principle the compression pump differs from the two-valve exhaust pump only in that the valves V and V', which are held in place by springs, operate downward instead of upward, Fig. 76. When the piston is pushed downward, valve V closes and the air in the cylinder C is forced into the storage tank, which is provided with a pressure gauge.

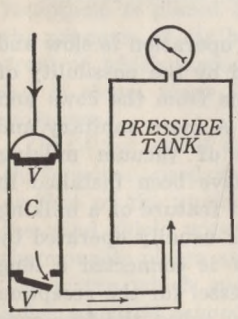


FIG. 76
Compression pump

One of the most familiar examples of an air compressor is that furnished by the common bicycle or automobile pump, an outline section of which is shown in Fig. 77. Connected with this pump there are two valves, one c, in the cylinder of the pump, and the other s in the tire, the latter being held in place by a spring. On the upward stroke of the piston P, valve

s in the tire closes and the cup-shaped piece of leather *c* allows the air to pass down to the chamber *C*. On the

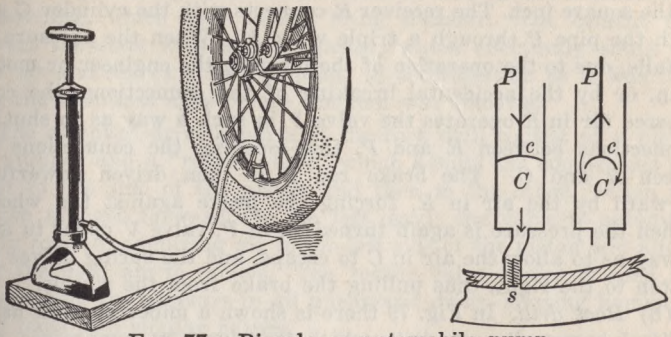


FIG. 77.—Bicycle or automobile pump

downward stroke of *P*, valve *c* closes and the air is driven into the tire. The pressure pumps used in garages are usually operated by electric motors.

87. Commercial and Industrial Uses of Compressed Air. Next to the application of heat and electricity in the arts and sciences, the applications of air pressure in the commercial and industrial affairs of life are probably more commonly met with than are those of any other phase of physical science. Air under pressure is used in garages for the pumping up of automobile tires. All modern railway cars, both passenger coaches and freight cars, are now provided with air brakes. Riveting hammers, rock drills, stone cutting tools, pneumatic jacks, and innumerable other devices are driven by air pressure.

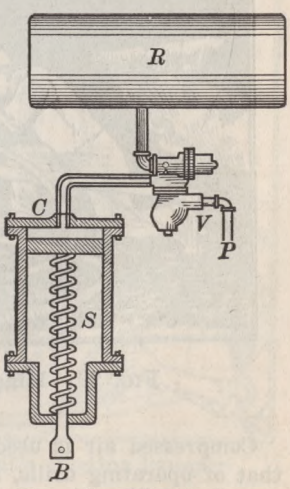


FIG. 78.—Westinghouse air brake

(a) *Westinghouse air brake.* One of the most important commercial applications of compressed air is that employed in the Westinghouse air brake, now used on nearly all steam and electric cars. The principle upon which the air brake works is illustrated in the diagram,

Fig. 78. A compression pump on the engine supplies air to the receiver R through the pipe P under a pressure of about 70 pounds to the square inch. The receiver R connects with the cylinder C and with the pipe P through a triple valve V. When the pressure on P falls, due to the operation of the lever by the engineer or motorman, or by the accidental breaking of the connections, the compressed air in R operates the valve V in such a way as to shut off connections between R and P, thus opening the connections between R and C. The brake rod B is then driven powerfully forward by the air in R, forcing the brake against the wheels. When the pressure is again turned on in P, valve V opens in such a way as to allow the air in C to escape, and the spring drives the piston to the right, thus pulling the brake from the wheels.

(b) *Rock drill.* In Fig. 79 there is shown a miner at work using a "jack-hammer" rock drill, which is driven by compressed air. With this machine one man can drill as many holes in an hour as he could in eighteen hours with the old-fashioned hand drill. Holes are drilled into the rock or ore, after which powder or dynamite is introduced into them and the rock blasted into relatively small pieces.



FIG. 79.—Miner using rock drill driven by compressed air

Compressed air is used in mines also for other purposes than that of operating drills, as, for example, the coating of the roof and walls with cement by means of a fine spray from a compressed air spray pump or "cement gun." One of the frequent causes of

mine cave-ins is the weathering of the slate of the roof and side walls. It gradually crumbles or scales off and suffers a consequent weakening, which may finally bring disaster. The cement gun covers the slate with a thin plaster, which effectually shuts out the air and leaves it as unexposed to deterioration as it was during the countless ages before the coal was removed.

The essential driving parts of the rock drill consist of a piston or plunger and a "rocker valve" which admits the compressed air first to one side of the piston and then to the other. When the piston plunges forward (to the left, say), the valve is automatically rotated, admitting air to the left side of the piston and allowing the used air to escape. The rocker valve device is used for alternating the pressure in all pneumatic drills, riveting hammers, and tools of a similar nature.

(c) *Riveting hammer.* In modern ship building and in the structural work of the "sky scrapers" of our cities, steel work is extensively used. Modern city buildings of any considerable size are today constructed almost exclusively of steel, concrete, brick, and tile. In the erection of buildings of this sort a steel framework forms the supporting part of the structure. The steel beams composing this framework are riveted firmly together, forming when completed practically a solid piece. The countless iron rivets required are driven firmly into place and fastened (riveted) by means of riveting hammers, which are driven by air pressure. On some jobs small hand riveters are used. On most large construction

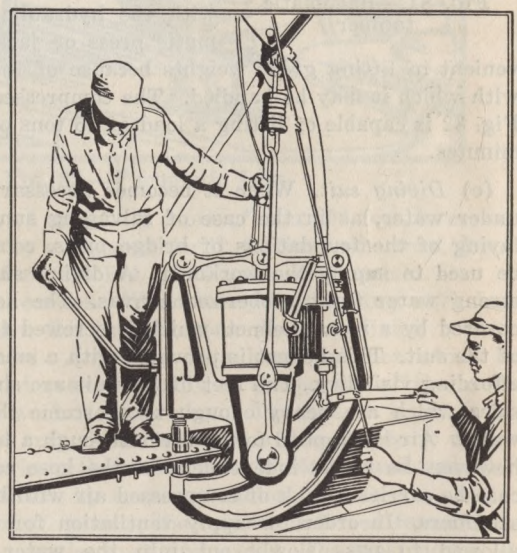


FIG. 80.—Pneumatic riveting hammer

operations, however, large power hammers are employed, Fig. 80. These large riveting hammers are supported by means of a traveling crane. They are usually operated by two men, one controlling the air supply and the other directing the hammer. As many as twenty-five rivets a minute have been set by such machines.



FIG. 81.—Pneumatic tamper

Compressed air tampers, Fig. 81, are now used to break up concrete pavements in street improvement work. The pneumatic tamper works on the same principle as the rock drill and the riveting hammer.

(d) *Pneumatic jack.* Air pressure in the pneumatic jack, Fig. 82, can be used for lifting great weights in a manner exactly similar to that employed in the use of water pressure with the hydraulic press. The pneumatic press or jack is extremely convenient in lifting great weights because of the comparative ease with which it may be handled. The compressed air jack shown in Fig. 82 is capable of lifting a load of 35 tons one foot high in two minutes.

(e) *Diving suit.* When it becomes necessary to perform work under water, as in the case of salvaging sunken vessels, in the laying of the foundations of bridge piers, compressed air has to be used to supply the workmen. A diving suit is made of very strong water tight rubberized canvas. The head of the diver is covered by a metal helmet, which is screwed tightly to the collar of the suit. This helmet is provided with a small circular window, affording visibility. The feet of the suit are shod with heavy lead soles which are heavy enough to overcome the buoyancy of the water. Air is supplied to the diver through a hose. In some cases, however, he works independent of the hose connection, in which case he carries a tank of compressed air with him, strapped to his shoulders. In order to supply ventilation for the suit, the air is allowed to pass slowly out into the water through a valve. This means that the air pressure within the suit must be slightly in excess of the water pressure under which he works. Communication is usually made between the diver and his assistant at

the surface by means of a telephone cable. In general divers do not work at a depth greater than 60 or 70 feet, which is considered

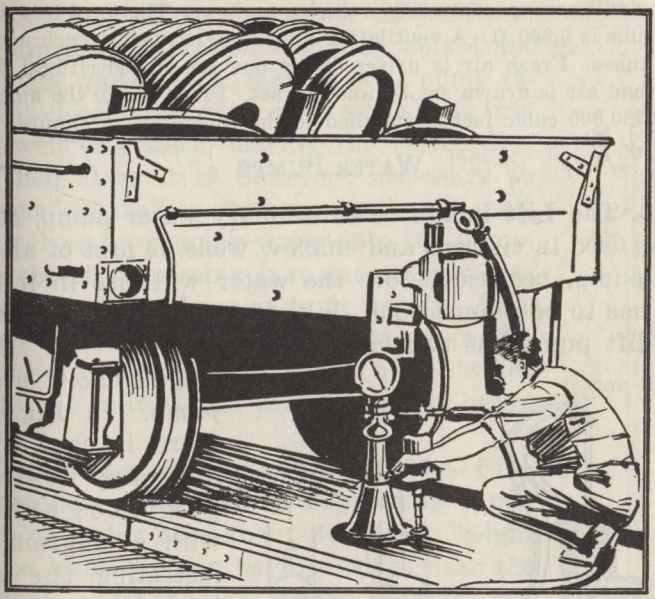


FIG. 82.—Pneumatic jack

to be about the limit of safety. A depth of over 300 feet, however, was reached in 1915 by a diver of the U. S. Navy who went down to investigate the submarine F-4, which sank off Honolulu.

(f) *Diving bell or caisson.* When heavy engineering work has to be carried on below the surface, some form of diving bell or caisson has to be used. A *caisson* is a large cylinder which is loaded at one end and sunk vertically to the bottom of the river or harbor where the work is undertaken. The caisson is closed at the top by means of trap doors, and air is forced into it under pressure sufficient to keep the water out. Workmen and material are lowered to the bottom by means of cables.

In the construction of tunnels, the caisson is forced forward in a horizontal direction as the work progresses. The construction of the Holland tunnel for vehicle traffic under the Hudson River between New York City and Jersey City was a triumph for compressed air, the use of which made it possible for the engineers to

force the great caissons through from end to end. This tunnel, which was seven years in building, cost $50,000,000. The twin tubes will accommodate 3800 vehicles an hour. The total length of the tube is 9,250 ft. A ventilating station is located at each end of the tubes. Fresh air is driven in by one set of centrifugal fans and bad air is drawn out by another set. Fresh air to the amount of 3,760,300 cubic feet is supplied to the tunnel every minute.

Water Pumps

88. The Lift Pump. The ordinary water pump, such as is used in cisterns and shallow wells, is first of all an air pump, because before the water will rise in it the air has to be pumped out. Like the two-valve air pump, the lift pump has two valves.

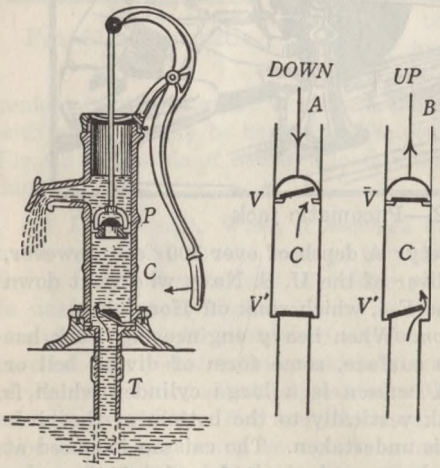


FIG. 83.—Lift pump

Shallow well pump. In Fig. 83 there is shown an ordinary cistern pump, together with a diagram illustrating the action of the valves on both the downward and upward strokes of the piston. The pump contains two valves, V and V', both of which open upwards. Diagram A illustrates the operation of the valves during the downward stroke; B shows the action of the valves during an upward stroke. When the pump is first operated, the action of the piston and valve exhausts the air from the chamber C. During this stage it acts as an air pump. The atmospheric pressure upon

the surface of the water in the well forces the water up into the pump until both valves become submerged. As the pump continues to operate the valves act exactly as explained in connection with diagrams *A* and *B*.

If the valves of a shallow well pump do not fit closely, the water may settle back into the well or cistern, leaving the valves dry. In such cases it is therefore necessary to "prime" a shallow well pump, that is, wet the valves by pouring water into the pump. The object of priming is to cause the valves to swell sufficiently so as to fit tightly. Also, when the valves have dirt in them or the hinges are not free, the weight of the water seats the valves, the air not being heavy enough.



FIG. 84
Deep-well pump

Deep-well pump. Since the pressure of the atmosphere is capable of supporting a column of water only about 34 feet in height at sea level, a pump having valves near the surface, as in the case of the ordinary cistern pump, would be of no value in lifting water from a well more than 30 to 34 feet deep. To make a deep-well pump effective, it is necessary to place both valves near the water. In the case of a deep well, such as is used in drive wells for example, both valves are usually placed within a few feet of the water, Fig. 84.

EXERCISES. **22.** To what height will water rise in a lift pump from which the air has been thoroughly exhausted (a) when the pressure upon the surface of the water is one atmosphere? (b) 72 cm?

23. Is it possible to raise brine, specific gravity 1.18, from a tank by means of a lift pump if the upper valve of the pump is 29 ft. above the surface of the brine? Explain.

24. Explain how it is possible to pump water out of a deep well, one for example that is 100 ft. deep.

89. The Force Pump. A *force pump* serves to lift the water from the well and also to deliver it in a steady stream under pressure. In the force pump there is no valve in the plunger *P*, Fig. 85. It, however, has two valves, as in the lift pump, one in the lower end of the cylinder and the other in the delivery tube. Both valves open upwards. The compression chamber is partly filled with air. When water is forced into this chamber by the downward stroke of the plunger *P*, the air in this chamber is compressed. During the upward stroke of *P*, the valve *V* closes and the air cushion expands, forcing the water out through the delivery pipe *D* in a steady stream.

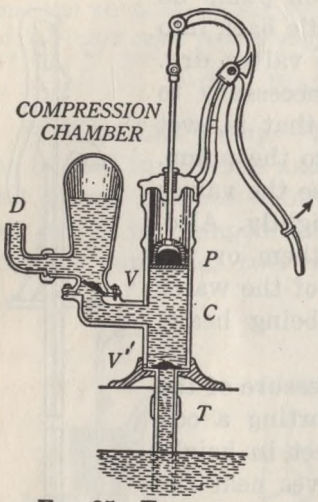


FIG. 85—Force pump

90. The Siphon. A *siphon* is a device for transferring liquids from a given level to a lower level, over an intervening elevation. It depends for its operation on atmospheric pressure. In Fig. 86 there is shown a siphon in operation. The short arm *ab* of the siphon is measured from the point of application (surface of the liquid) of atmospheric pressure in vessel *A* to the highest point of the siphon; the long arm *ce* is measured from the highest point of the tube to the level at which the liquid emerges from the tube.

To start a siphon working, place one end (the end marked *A*, Fig. 86) in a dish of water and by suction or

otherwise fill the tube with water. Place the finger over the end of the long arm *B* to keep the water from running out while the tube is being placed in position. On removing the finger from the end of *B*, water will at once begin to flow. The liquid in a siphon always flows from the short arm to the long arm. The student should remember that by short arm we mean the distance *ab*, and by long arm the distance *ce*.

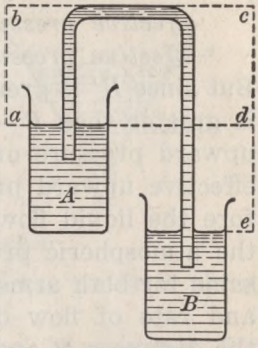


FIG. 86.—The siphon

Since the operation of a siphon depends upon atmospheric pressure, it follows that a liquid can be siphoned over an elevation as high as the atmospheric pressure will support a column of the liquid in question. We have learned that at sea level atmospheric pressure will support a column of water 34 feet in height, and a column of mercury 76 centimeters in height. This means that it would be impossible to siphon water over an elevation greater than 34 feet or mercury over an elevation greater than 76 centimeters.

91. Action of the Siphon Explained. We have learned (Art. 36) that the pressure of a liquid at any given depth is equal to *HD*, where *H* is the distance from the level in question to the surface, and *D* is the density of the liquid. Now in each arm of the siphon there are two pressures to be taken into account, one the atmospheric pressure acting upward on both arms, and the other the pressure due to the weight of the liquid acting downward. The difference between these two pressures in the arms of the siphon is called the *effective pressure*. Let *P* be the pressure of the atmosphere acting upward on the liquid in both arms, *HD* the downward pressure

of the water in the short arm, and $H'D$ the downward pressure of the water in the long arm. Then

effective pressure, short arm, $= P - HD$, and

effective pressure, long arm, $= P - H'D$

But since H' is greater than H, it follows that $P - HD$ is greater than $P - H'D$. In other words, the effective upward pressure on the short arm is greater than the effective upward pressure on the long arm, and therefore the liquid flows toward the long arm. Also, since the atmospheric pressure P and the density D are the same for both arms, the whole question of the direction and rate of flow depends upon the relative values of the distances H and H'.

EXERCISES. **25.** (a) When the barometer reads 76 cm, to what height in feet will the atmospheric pressure support a column of water? (b) Under these conditions of atmospheric pressure, can water be siphoned over an elevation of 35 ft.? Explain.

26. What is the maximum elevation in feet over which sulphuric acid can be siphoned at sea level, the density of sulphuric acid being 1.84 g/cm^3?

27. What will happen to the liquid in the siphon if the two arms are of the same length, that is $H = H'$?

28. In which direction will the liquid flow if the siphon tube, Fig. 86, be cut off between c and d? Explain.

29. What is the effect on the rate of flow of increasing the length of the long arm, H'?

30. Will a siphon work under a bell jar from which the air has been exhausted? Explain.

92. The Intermittent Siphon. In Fig. 87, at the left, there is shown a section of a spring which operates on the principle of the siphon. The flow from such a spring is intermittent, and hence it is known as an "intermittent spring." The principle upon which it works may be demonstrated by means of an intermittent siphon. Seal a bent tube into a funnel, as shown in Fig. 87, right. Pour water slowly into the vessel. When the water rises to the top of the tube a, the siphon begins to operate, and

continues to flow until the water falls below the opening

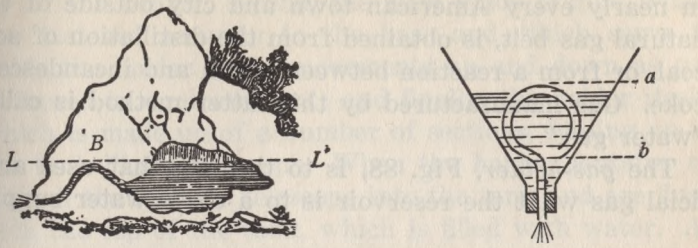


FIG. 87.—Intermittent siphon

of *b*. The siphon ceases to flow until the water is above *a*, when it starts again.

EXERCISES. **31.** What must be the shape of the outlet tube, Fig. 87, in order that the flow from the spring shall be intermittent?

32. (a) How high must the water rise in the spring before the flow commences? (b) When the flow once starts, how long will it continue?

GAS SUPPLY

93. Natural Gas. The supply of gas to many American cities for heating and cooking purposes is of almost as great importance as that of the supply of electricity. Natural gas occurs in certain parts of Pennsylvania, Ohio, Indiana, Oklahoma, Texas, and other states in connection with coal oil deposits. When the gas "pockets" or underground reservoirs are tapped, the natural gas, which is under very high pressure, sometimes rushes out with enormous force. In the early days, immense quantities of gas were allowed to go to waste, in many cases being burned as it flowed out from the well.

Natural gas is now piped from the wells to service lines and is thence distributed to consumers, or sent to stills where by-products are condensed from it. The supply of natural gas, however, is rapidly diminishing. Indeed in many places where a few years ago it was abundant the supply has entirely given out.

94. Artificial Gas. Artificial gas, which is now used in nearly every American town and city outside of the natural gas belt, is obtained from the distillation of soft coal, or from a reaction between steam and incandescent coke. Gas manufactured by this latter method is called "water gas."

The *gas-holder*, Fig. 88, is to the city that uses artificial gas what the reservoir is to a city's water supply.

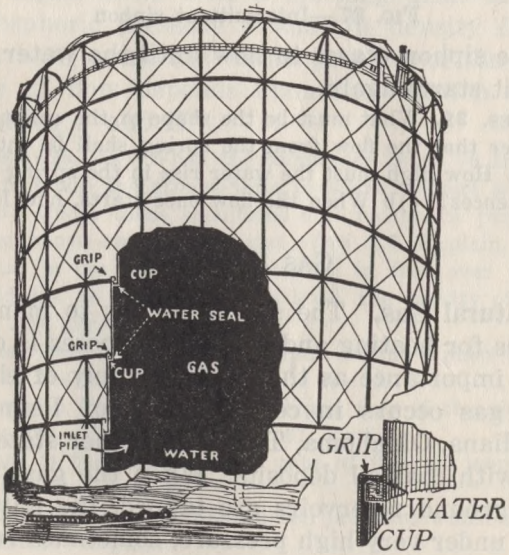


FIG. 88.—Gas holder

Instead of being located on a hill as is the water reservoir, the gas reservoir is usually located at the lowest point. Water flows down to the consumers, while gas, which is lighter than air, flows up. A gas tank is made up of three elements: First, the concrete foundation which is in the form of a circular reservoir extending below the surface of the ground and which contains

several million gallons of water, depending on the capacity of the holder; second, the steel guide frames which are fastened securely to the base and which serve to guide the holder in its movements up and down as gas is forced in or drawn out; and finally, the holder itself, which is made up of a number of sections, folding up in the tank like a telescope. When the holder contains no gas, all the sections telescope into the tank and are flush with the top of the tank, which is filled with water. As gas is forced into the holder through the inlet pipe, which passes up through the water in the tank, the first or inner section, which carries the roof or crown, is inflated; and when it reaches the top of the tank, it picks up the second section by means of what is known as the cup. This cup runs around the whole outside circumference of the holder, forming the bottom edge of each section, just as if the edges were turned over and bent upward to hold water. As each section rises it picks up the grip or inverted cup of the next section, and as each cup is filled with water there is formed what is known as the "water lute," or seal, which prevents the gas from escaping.

When the holder is fully inflated, all the sections are lifted with their respective cups filled with water. The lowest edge of the lowest section is considerably below the surface of the water in the tank. This prevents the gas from escaping at the bottom. The total floating weight of the holder, the weight which the gas is compelled to lift, may be as great as several hundred tons. This weight presses down on the gas contained in the holder to the extent of a little less than half a pound to the square inch.

95. The Gas Meter. The *gas meter* is a device which operates under the pressure of gas in the city mains, and thus registers automatically in cubic feet the quantity

of gas passing through it. It consists essentially of a double bellows, as shown in Fig. 89. Gas enters through the pipe *P* and passes down into chambers *A* and *C*, pushing the bellows to the right. This forces the gas in chambers *B* and *D* out through the delivery pipe *Q*. Now the motion of the bellows to the right causes the valve *V*

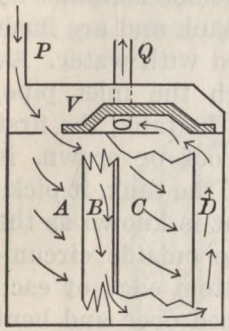


FIG. 89

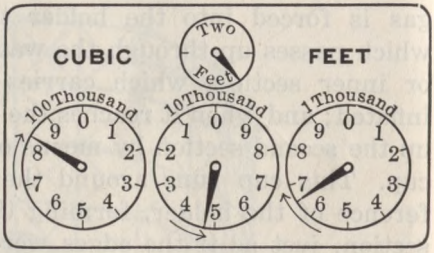


FIG. 90

Gas meter, and recording dials

to move to the left, thus closing the opening into *A* and opening that into *D*. The gas now flows in on the right of valve *V*, and the walls of the bellows move to the left, pushing the gas in *A* and *C* out through the pipe *Q*. When the bellows is moved to the left, the valve *V* again slides to the right and the process is repeated. The valve *V* connects not only with the bellows, but also with a train of wheels which registers on a set of dials the number of cubic feet of gas consumed.

How to read the gas meter. The small top dial, Fig. 90, which is marked "Two Feet" is called the "testing circle" and is used in testing the meter. The indication of the hand of the testing circle is ignored in the ordinary reading of the meter. Note that the hand of the 100 thousand dial and that of the 1 thousand dial rotate in a clockwise sense, while the hand on the middle dial rotates in the

opposite sense. To read the meter, begin at the left hand dial and read to the right, noting on each dial the number which was *last passed over* by the hand. Thus, in Fig. 90, the readings are:

100 thousand dial.....80,000 cubic feet
10 thousand dial..... 4,000 cubic feet
1 thousand dial..... 600 cubic feet

That is, the total number of cubic feet registered by the meter is 84,600.

Buoyancy of Air

96. Buoyancy of the Air. According to Archimedes' principle, bodies immersed in a fluid are buoyed up by a force equal to the weight of the fluid displaced. Now air is a fluid and it also possesses weight; hence all bodies in air are buoyed up by a force equal to the weight of the air displaced. The true weight of a body is its weight in a vacuum. In many accurate scientific measurements of mass it is necessary therefore to reduce the values obtained in air to those corresponding to a vacuum.

There is an old saying that "a pound of feathers is heavier than a pound of lead." This is in a sense true, because of the buoyancy of the air, as can be shown by the following experiment: Counterbalance in air a hollow spherical vessel against a small solid metal sphere, Fig. 91. Of course, the buoyancy of air upon the large sphere is greater than that upon the small solid sphere. Now place the balance under the receiver of an air pump and exhaust the air. As soon as the air is withdrawn from the receiver and its buoyant effects removed, the balance is no longer in equilibrium, the larger sphere now overbalancing the small one. Thus the two

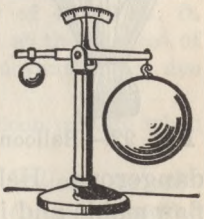


Fig. 91
Buoyancy of air

bodies which apparently had equal masses in air are found to have different masses in a vacuum. The large hollow sphere corresponds to the pound of feathers and the small solid sphere to the pound of lead.

To get the true weight of any given body, it is necessary to make a correction to the observed weight for the buoyancy of the air. This correction, however, is necessary only when very accurate values are desired.

Example. A given body having a volume of 1250 cm^3 weighs 980 g in air. What is its weight, reduced to a vacuum? *Solution*: Since the volume is 1250 cm^3 and since a cubic centimeter of air under standard conditions weighs 0.00129 g, it follows that the body in question is buoyed up by a force equal to $1250 \times 0.00129 = 1.61$ g. The weight of the body then referred to a vacuum is $980 + 1.61 = 981.61$ g.

97. The Principle of Buoyancy Applied to the Balloon. A *balloon* is an air-tight spherical bag, usually made of rubberized cotton or silk, filled with a gas lighter than air, Fig. 92. The buoyant force tending to lift the balloon is equal to the weight of air which it displaces. The gas generally used for inflating balloons is either hydrogen or helium. For exhibition flights illuminating gas is sometimes used. Hydrogen is the lightest known gas suitable for buoyancy purposes. Because of its inflammability when mixed with air, it is very dangerous. Helium gas, on the other hand, is non-inflammable and is therefore coming into general use. Previous to 1915 very little helium gas had been isolated. Its cost was about $2000 per cubic foot. During the World War, the U. S. Government discovered that helium

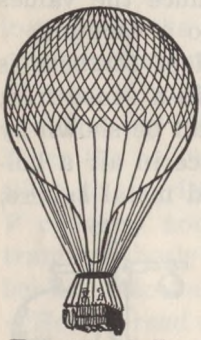


FIG. 92.—Balloon

could be obtained from our Southwestern oil and gas wells. The U. S. Helium Plant at Fort Worth, Texas, now produces helium at a cost of about 2 cents per cubic foot.

The relative densities of gases used for buoyancy purposes are as follows:

Air	= 0.08	lb. per cu. ft.
Illuminating gas	= 0.05	lb. per cu. ft.
Helium	= 0.0112	lb. per cu. ft.
Hydrogen	= 0.006	lb. per cu. ft.

Example. Given two toy balloons of the same volume, one containing a cubic foot of hydrogen and the other a cubic foot of illuminating gas, to find the net buoyant effect on each, neglecting the weight of the balloon in each case. *Solution*: The weight of air displaced by each balloon is the same; namely, 0.08 pound. The net buoyant effect in each case will therefore be the difference between the weight of air displaced and the weight of the balloon; that is, (a) for hydrogen the net buoyant effect will be $0.08 - 0.006 = 0.074$ pound per cubic foot; (b) for illuminating gas, $0.08 - 0.05 = 0.03$ pound per cubic foot. Thus we see that the net buoyant force acting upon the hydrogen balloon is about 2.5 times as great as that acting upon a balloon containing illuminating gas.

EXERCISES. **33.** The balloon in which Captain Gray made his record altitude flight (Art. 63) had a capacity of 42,470 cu. ft. Suppose that the balloon had been fully inflated at the surface of the earth, what would have been the lifting (buoyant) force due to the displaced air?

34. What weight of hydrogen gas did the balloon contain when fully inflated?

98. The Dirigible Balloon. A *dirigible balloon* or *airship* consists of a cigar-shaped receptacle for holding the gas, below which is suspended the car for the accommodation of the driving engine and the passengers, Fig. 93. Propulsion is obtained by means of the rotation of fan-

like propellers. The modern dirigible is of enormous size and is provided with very powerful engines. The British dirigible R-34, which was the first to cross the

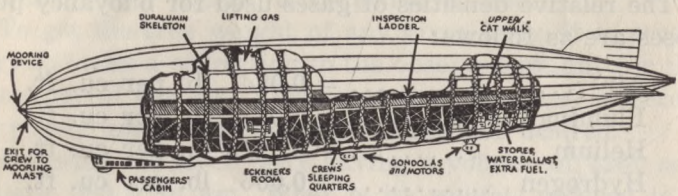


FIG. 93.—Dirigible balloon

Atlantic, arriving in America (Mineola, Long Island) July 6, 1919, had a capacity of 2,000,000 cubic feet, and was provided with five engines giving a total of about 1300 horsepower. The trip was made from Scotland to New York in 108 hours; the return trip was made in 75 hours.

The first circumnavigation of the globe was accomplished by an expedition under the command of Magellan, who set sail from Spain in 1522. The entire journey took approximately 3 years. In 1929 Captain Hugo Eckener, in command of the German dirigible, *Graf Zeppelin*, completed a trip around the world, including stops en-route, in 21 days, 8 hours, and 26 minutes.

stop

SUMMARY

1. Since gases, unlike liquids, are highly compressible, the pressure depth law does not apply. In liquids pressure is proportional to depth; in gases this is not true.

2. A pressure of one atmosphere is equivalent to

$$76 \text{ cm of mercury} = 1033.3 \text{ g/cm}^2$$
$$30 \text{ in. of mercury} = 14.7 \text{ lb./sq. in.}$$

3. The direction of winds at a given place on the earth's surface is determined by difference of air pressure (barometric variation) and the rotation of the earth.

4. When the velocity of a fluid is increased at a given point, the lateral pressure at that point is diminished.

5. Pascal's law: Pressure applied to an enclosed gas is transmitted equally in all directions.

6. Boyle's law: The temperature remaining constant, the volume of an enclosed gas is inversely proportional to the pressure. When the volume of a gas is decreased, its pressure and density increase. The equation for Boyle's law is $P\ V = C$.

7. An exhaust air pump works on the principle that the air in the receiver tends to expand as exhaustion continues.

8. A water pump of the lift type depends on atmospheric pressure for its effective operation. Water may be lifted by such a pump to a height equal to that which the atmospheric pressure will support a column of water.

9. A siphon depends upon atmospheric pressure for its operation. The effective upward pressure is

$$\textit{effective pressure, short arm,} = P - HD$$
$$\textit{effective pressure, long arm,} = P - H'D$$

Since H' is greater than H, the liquid always flows from the short arm toward the long arm.

10. Archimedes' principle of buoyancy applies to gases as well as to liquids. A body in air is buoyed up by a force equal to the weight of the air displaced.

11. The effective lifting force on a balloon is equal to the weight of the air displaced minus the weight of the balloon and its enclosed gas.

REVIEW EXERCISES

A

1. (a) What are the principal gases of which the atmosphere is composed? (b) What is their ratio by volume?

2. State briefly the three methods employed to investigate the density of the atmosphere.

3. What is the density of the air at the top of Mt. Blanc as compared with that at the earth's surface?

4. Explain how we may prove that air has weight.

5. Under standard conditions (a) what is the weight of one cm^3 of air? (b) one cu. ft.?

6. (a) Describe briefly Torricelli's experiment. (b) What did he prove by this experiment?

7. Describe von Güricke's Magdeburg hemisphere experiment.

8. Explain the construction and use of a barometer.

9. A pressure of one atmosphere is equivalent to (a) how many g per cm^2? (b) lb. per sq. in.?

10. A pressure of one atmosphere is equivalent to how many cm of mercury? (b) inches of mercury?

11. Describe briefly the aneroid barometer.

12. (a) What advantage has an aneroid barometer over a mercury barometer? (b) Which is the more liable to get out of order?

13. A pressure of 998 g per cm^2 is equivalent to what barometric reading, in cm?

14. A laboratory barometer at a given time reads 74.87 cm. When carried to a room 3 floors above, it reads 74.75 cm. What is the vertical distance between the two positions, in ft.?

15. Explain the use of the barometer as a weather indicator.

16. Explain the circulatory motion of the air in regions of low atmospheric pressure.

17. State Boyle's law, and explain the meaning of each term in the equation $P\ V = C = P'\ V'$.

18. A given mass of air under pressure of 2000 g/cm^2 has a volume of 300 cm^3. If the pressure be increased to 3000 g/cm^2, what will be its volume, the temperature remaining constant?

19. How will the density of the air (Ex. 18) be affected by the increased pressure, and how much?

20. Make a sectional drawing of a two-valve air pump, and explain the operation of the valves.

21. Make a sectional drawing of a water jet pump or aspirator and explain.

22. Wherein does the compression pump differ from an exhaust pump?

23. Why is air under pressure supplied to divers and men working in diving suits and caissons?

24. When the bubbles of air escape from the valves of a diver's suit, they grow larger as they rise. Why?

25. Make a sectional drawing of a lift pump and explain the operation of the valves.

26. Make a sketch of a force pump and explain. What is the air dome or compression chamber for?

27. By means of a sketch explain the operation of a siphon.

28. Why does the liquid always flow toward the long arm?

29. Explain why a liquid cannot be siphoned from a given level to a higher level.

30. Why does a strong wind blowing across a chimney create an upward draft?

31. Make a sectional drawing of an atomizer and explain.

B

32. Make a sketch to illustrate the working parts of a Bourdon gauge.

33. What is the difference between gauge pressure and absolute pressure?

34. Make a sectional drawing of the mercury air pump and explain.

35. A force pump delivers water into a tank. The distance between the piston and the surface of the water in the tank is 20 ft. What pressure (atmospheric pressure + water pressure) is exerted by the piston?

36. A concrete tank 6 ft. in height rests upon the ground with its floor somewhat above the ground level. The tank is filled with water, the only opening being in the top. Explain how the liquid in this tank may be emptied by the use of a rubber hose. How long would the hose have to be?

37. What is the greatest height over which gasoline can be siphoned when the barometer reads 75 cm?

38. How many cu. ft. of gas does the meter shown in Fig. 93*a* register?

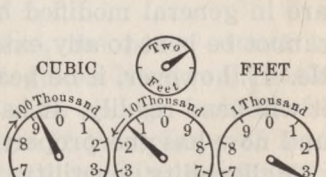


FIG. 93*a*

C

39. Describe briefly the mercury vapor pump. What are the three essential parts? What vacuum may be obtained by this pump?

40. A sphere having a volume of 100 cm^3 weighs one kilogram in air. What is its weight in a vacuum?

41. A block containing 2 cu. ft. weighs 100 lb. in air. What is its weight referred to a vacuum?

42. A balloon has a volume of 60,000 cu. ft. It is filled with helium gas. Find (a) the weight of the helium in the balloon; (b) the buoyant force on the balloon due to the weight of the air displaced.

NOTE: For additional problems see Appendix.

CHAPTER IV

MOLECULAR MECHANICS

99. Introductory. The properties of matter may be divided into two general classes, general and special. *General properties* of matter are those which are true for all substances, such properties for example, as extension, indestructibility, and weight. *Special properties,* on the other hand, are those which are characteristic of certain substances only for given conditions. For example, *hardness* is a special property of the diamond, *tenacity* of iron, and *brittleness* of glass at ordinary temperatures. There are a great many such special properties. In general these properties are due to the nature of the forces which act between the molecules. These molecular forces are in general modified by heat. A glass rod when cold cannot be bent to any extent without breaking; it is brittle. If, however, it be heated in a flame for few minutes, it will bend readily. It has lost its property of brittleness and now has the property of flexibility.

Malleability, ductility, and tenacity are special properties of certain substances. *Malleability* is that property by virtue of which certain metals may be beaten out into thin sheets. A good example of matter in a malleable condition is the tin foil which is wrapped around many articles displayed for sale in stores. The foil is made by hammering or rolling the metal tin out into very thin sheets. Gold is one of the most malleable of metals. *Ductility* is that property by virtue of which a substance may be drawn out into threadlike strands, or, in the case of metals, into wire. Glass at ordinary tem-

peratures is very brittle, but when heated it becomes very ductile. If a small glass tube be placed in a flame for a few moments, it becomes soft and pliable. It can then be drawn out into very fine threads which are flexible and highly elastic. Most metals are ductile and hence are capable of being drawn out into wire. Platinum is one of the most ductile metals known. It can be drawn into wire so fine as to be almost invisible. *Tenacity* is that property of a substance by virtue of which it resists being torn apart. Paper is quite easily torn; its tenacity is low. Iron, on the other hand, is not easily torn apart; its tenacity is high.

100. Cohesion and Adhesion. The attraction which exists between molecules of the same kind is called *cohesion.* Attraction between molecules of unlike kind is called *adhesion.* The particles of a piece of chalk are held together by cohesion. On the other hand, when crayon is drawn across the blackboard, a chalk mark is made; the attraction between the chalk and the board is due to adhesion.

If a piece of chalk be broken, the parts cannot be united again by pressing them together because the molecules cannot be brought into close enough contact for the molecular forces to operate. If two pieces of lead, however, are forced together by a twisting motion, the parts will cohere.

A broken plate can be mended by means of cement, an adhesive substance in a liquid condition, which permits of the molecules being brought into intimate contact. The blacksmith unites two pieces of iron by first heating them red hot and then welding (hammering) them together. The hammering is done for two purposes, (a) to force the molecules into intimate contact, and (b) to give the iron the desired shape.

If a clean glass rod be thrust into water and then

withdrawn it will be found that a drop of water clings to the glass. This means that the attraction of water for glass (adhesion) is greater than the attraction of water for water (cohesion).

101. Viscosity. *Viscosity* is that property of matter in a fluid or semi-fluid state which tends to retard its tendency to flow. We are all familiar with the fact that some fluids flow more readily than others. Alcohol, gasoline, and water flow freely; their viscosity is relatively low. Heavy oils, syrup, and molasses, on the other hand, do not flow freely; their viscosity is high. In cold weather syrups and oils become less fluid and consequently the friction between their molecules increases. For this reason the viscosity of a fluid in general increases as the temperature falls.

Viscosity is due to the friction which is encountered by the molecules of a fluid as they slide one over the other; that is to say, viscosity is the molecular friction in fluids which tends to retard their flow. All fluids possess viscosity to some degree. The viscosity of liquids is relatively high as compared with that of gases.

EXERCISES. **1.** Why is the viscosity of a liquid greater than that of a gas?

2. The oils used in an automobile engine are variously designated as light, medium, and heavy. How do the viscosities of these three grades of oil compare?

3. Why does heating an oil diminish its viscosity?

MOLECULAR FORCES IN GASES

102. The Kinetic Theory of Gases. The molecules of a substance, whether in the solid, liquid, or gaseous state, are not at rest with respect to each other but are in constant motion, bounding back and forth as they strike one upon the other or against the sides of the containing vessel. The speed of the molecules of a substance in a

gaseous state is enormously greater than that of liquids or solids. It is estimated that the speed of air molecules under ordinary conditions is about 1460 feet per second, while that of the hydrogen molecule is nearly 6000 feet per second.

The fact that the molecules of which matter is composed are in constant motion and that in the case of some gases (hydrogen) their velocity may reach the enormous value of more than a mile a second explains many of the phenomena of the diffusion, expansion, and compression with which we are familiar. We shall now consider, first, some of the experimental evidence which we have for the foregoing statements, and second, some of the applications of this kinetic theory, especially as relating to gases.

One of the most characteristic properties of a gas is its tendency to expand indefinitely. A confined gas always completely fills the containing vessel. Thus if a little gas having a strong odor be introduced into a room, it penetrates to all parts of the room in a very short time, as can be determined by its effect upon the organs of smell. And this is true even if the gas in question is much heavier than the air in the room.

One of the best evidences of molecular motion is that furnished by the diffusion of two fluids. If a tube containing hydrogen (atomic weight 1) be inverted over a similar tube containing oxygen (atomic weight 16) the two gases will mix, notwithstanding the fact that the lower gas (oxygen) is 16 times as heavy as the upper. The diffusion of one gas into another is due to the motion of the molecules of the gases. Hydrogen and oxygen form an explosive mixture. The fact that diffusion actually takes place in the case of hydrogen and oxygen may be demonstrated by applying a lighted match to either tube after the two gases have been in contact for some time.

A slight explosion occurs, thus showing that the gases have mixed.

Experiments of the sort recorded above, and many others like them, make very clear the fact that the molecules of a gas are not at rest even when the gas is apparently perfectly still, but are in constant and violent action.

103. Molecular Motion and Pressure. When we consider the constant motion and the great velocity of the molecules of a gas we are prepared to understand something of the cause of gas pressure which is exerted upon the walls of the containing vessel. The blow delivered by any one molecule is of course insignificant, but when these blows are delivered thousands of times a second by hundreds of thousands of molecules the result becomes relatively very great. The force exerted by a gas upon unit area depends upon two factors, as has just been suggested. These factors are, first, the speed of the molecule, and second, the number of molecules per unit volume. It has recently been determined that the number of molecules per cubic centimeter in a gas at atmospheric pressure and 0° C is 27 billion billion. While such a number is far beyond our comprehension, nevertheless the very magnitude expressed by these figures serves to emphasize the fact that molecular motion and gas pressure are closely related phenomena. We are also enabled to get a new insight into the significance of Boyle's Law. In Art. 76 we learned that, according to Boyle's Law, when the volume of a confined gas is decreased the pressure which it exerts is increased. That is, to use a concrete case, if the volume is diminished by one-half, the pressure is doubled, the temperature remaining constant. This is so because, when the gas is crowded into half the space which it formerly occupied, there are twice as

many molecules per unit volume and hence twice as many blows struck per second.

104. Diffusion of Gases through Solids. When the flying molecules of a gas strike against a porous solid, as an unglazed battery jar, we can readily understand that some of the particles will pass into or quite through the walls of the jar. If the gas in the vessel is air at the same pressure as that of the outside, the number of molecules that pass out will just equal the number that pass in, and as a result, the pressure on both sides will remain the same. But if the gas on one side of the wall is lighter and therefore has a greater speed than the gas on the other side, it follows that the passage of the molecules of the light gas in one direction will be in excess of the passage of the heavy molecules in the opposite direction, and consequently a difference of pressure will be created.

The passage of a relatively light gas through the walls of a porous vessel may be demonstrated by the following simple experiment. Lower a large glass jar filled with some light gas, such as hydrogen or illuminating gas, over a porous cup, to the lower end of which is sealed a bent glass tube containing colored fluid, Fig. 94. The light gas in the jar diffuses readily through the walls of the porous cup, thus giving rise to a pressure which is manifested by the rise of the colored liquid in the bent tube. If now the glass jar be removed from over the porous cup, the light gas which has diffused into the cup will now diffuse outward, thus creating a partial vacuum within the cup. This is shown by the fact that the colored liquid in the bent tube rises toward the cup, above its position of equilibrium *ab*.

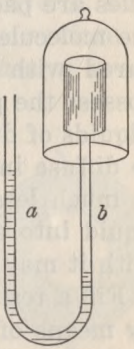


FIG. 94

The lighter a gas, the more readily will it diffuse

through porous solids. The rate of diffusion of two gases through a porous partition is inversely proportional to the square root of their densities. For example, the weight of hydrogen is to that of oxygen as the ratio of 1 : 16. Hence the rate of diffusion of hydrogen is to the rate of diffusion of oxygen as $\sqrt{16} : \sqrt{1} = 4 : 1$. That is, hydrogen will diffuse 4 times as fast as oxygen.

Some gases will diffuse through metals under certain conditions. Carbon monoxide (CO), a very poisonous gas, will diffuse readily through cast iron when the latter is red-hot. In this way carbon monoxide poisoning may arise from over-heated stoves in poorly ventilated rooms. Hydrogen, also, will diffuse through red-hot platinum.

Molecular Forces in Liquids

105. Liquids and Gases Compared. Liquids and gases are both fluids; they both possess the property of flowing. Liquids, however, differ from gases in several important particulars. Liquids always possess a definite surface; gases do not. Liquids are denser than gases; their molecules are packed relatively close together. The motion of the molecules of a liquid are therefore impeded, as compared with the free motion of gases. Liquids do not possess the property of unlimited expansion as do gases. Liquids of different densities when placed in contact tend to diffuse into each other, although the rate of diffusion is much less than that of gases. Diffusion of a heavy liquid into a lighter liquid placed above and in contact with it may be shown by the following experiment.

Fill a test tube two-thirds full of litmus solution. Now by means of a thistle tube pour some sulphuric acid into the tube below the litmus solution. The sulphuric acid is considerably heavier than the litmus solution, and the line of demarcation between the two is sharply defined. In a few hours this line of separation will have moved upward,

showing that the heavy acid is diffusing upward into the lighter liquid above it.

106. Surface Tension. One of the most striking illustrations of the action of molecular forces is found in the study of the surface of a liquid. Consider the molecules of a liquid, as shown in Fig. 95. Molecule *a*, for example, is attracted equally in all directions within the range of the molecular forces (represented by the circle), and hence is in equilibrium. A molecule in the surface layer *b*, for instance is not attracted equally in all directions, the force downward being greater than the force upward, because the molecules of water attract water with a greater force than do molecules of air. Hence it follows that the molecules of the surface layer are drawn downward, and are therefore a little closer to their neighbors than are the molecules of any other part of the liquid. In other words the molecules of the surface layer are packed more closely together than are the molecules of the lower layers. This gives rise to a tension in and parallel to the surface of the liquid, called *surface tension. The surface of every liquid is under a state of tension and acts exactly as if it were a stretched membrane.*

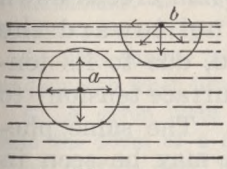
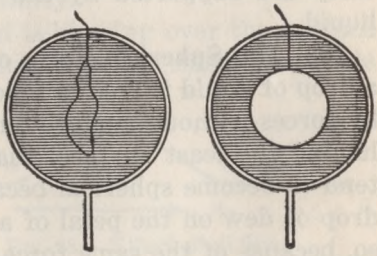


FIG. 95

107. Illustrations of Surface Tension. (a) If a wire ring having a loop of thread tied to it, be dipped into a soap solution and then withdrawn so that a film is formed across it, the loop of


FIG. 96.—Surface tension pulls the loop out into a circle

thread will lie on the face of the film. If now the film be punctured inside the loop, the thread will spring out into a circle, the tension of the film pulling the loop outward equally in all directions, Fig. 96.

(b) If a film be formed across the mouth of a funnel by dipping it into a soap solution, it will be observed that the film creeps down the funnel toward the small end, due to its tendency (surface tension) to contract. The same phenomenon may be seen in the case of the contraction of a soap bubble blown from a pipe, Fig. 97.

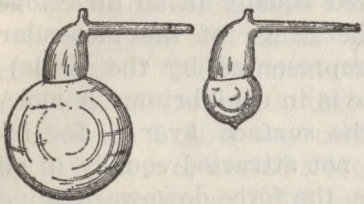


FIG. 97.—Surface tension causes the soap bubble to contract

(c) Let a film of water be spread out over a clean glass plate. Now put a drop of alcohol on the water and the film will be torn asunder at the point of contact, due to the fact that the surface tension of water is greater than that of alcohol.

(d) If by means of a wire holder a needle be carefully placed upon water, the membrane-like surface will cause it to float. Certain insects may be seen in the summer time running about on the surface of still water. They are supported by the tough surface layer of the liquid.

108. The Spherical Form of a Drop of Liquid. When a drop of liquid is free to follow its tendencies, unaffected by forces without, surface tension draws it into the form having the least surface, that is, a sphere. Rain drops tend to become spherical because of surface tension; the drop of dew on the petal of a rose is spherical, or nearly so, because of the same force. Drops of molten lead falling from a great height, as from the top of a shot tower, take on a spherical form as they fall due to surface

tension; and, cooling as they descend, retain their shape when they plunge into water at the bottom of the tower. Small globules of mercury are very nearly spherical despite the great weight of the metal, because of the tendency of the surface layer to force the liquid into the smallest possible volume.

EXERCISES. **4.** Why does a stream of water from a hose tend to break up into drops?

5. The sharp, jagged edge of a glass tube becomes rounded and smooth when heated in the flame of a Bunsen burner. Explain.

6. A given mass of a soft moulding clay is compressed first into the form of a cube and then into a sphere. Which has the least surface area, the cube or the sphere? Why do liquid drops always tend to assume the spherical form?

109. Oil on Water. If a drop of light oil be placed on water, it quickly spreads out over the surface. This is due to the fact that the surface tension of water is greater than that of the oil, and hence the latter is drawn out into a thin film.

The effect of oil in stilling a rough sea has been known from the earliest times. This method of calming a heavy sea about a ship is now frequently practiced. The oil makes the water calm only in the sense that it reduces the short choppy waves to a more or less smooth swell. The action of oil, therefore, in "stilling troubled waters" is not altogether a poetic fancy.

Let us assume that a wind is blowing over the surface of water, tending to heap it into waves. Consider a single wave, Fig. 98, moving in the direction from A to C and just entering the region BC, which is covered with oil. Here two surface tensions come into play: first, the surface tension of the

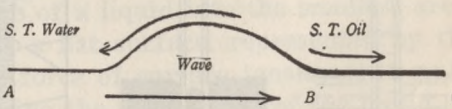


FIG. 98.—Effect of oil film on wave motion

oil, acting from *B* toward *C;* and second, the surface tension of the water acting in the opposite direction from *B* toward *A*. Now the surface tension of water is more than twice as great as that of oil, and hence the pull back toward *A* is greater than the pull forward toward *C*. The wave is therefore retarded and tends to flatten out. Thus the oil on the surface tends to prevent the heaping up of the water. It is these heaps of water which under wind pressure develop into the waves which are characteristic of high seas.

For surface tensions of different liquids, see Table VIII, Appendix.

110. Capillary Action. A capillary tube is one having a very fine, hairlike bore. If one end of a capillary tube of glass be thrust into water, the liquid rises, Fig. 99 *I;* when thrust into mercury, the liquid falls, Fig. 99 *II*. The elevation or depression of the liquid depends on the nature of the curved surface of the liquid within the tube. When the surface is curved downward, that is, depressed in the middle as in water, the

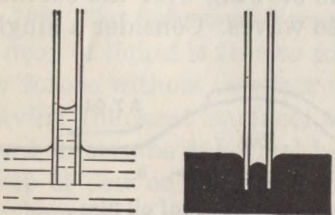


FIG. 99

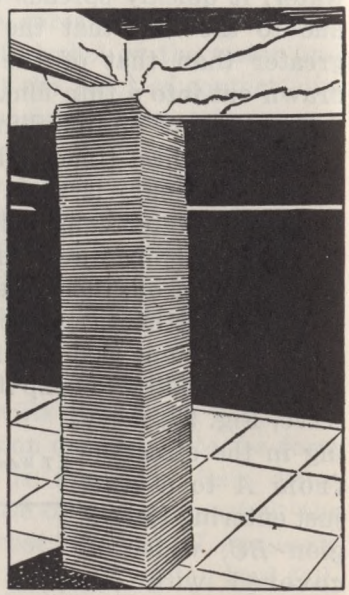


FIG. 100.—Heavy concrete beam cracked by the swelling of a pile of water-soaked boards

liquid rises; when curved upward as in mercury, it falls.

The rise of oil in the wick of a lamp, the passage of ink into a blotter, and the flow of ink in the slit of a pen are all familiar illustrations of capillary action. Capillary action accounts in the main for the passage of water into a rope, or between the fibers of wood, thus causing the wetted rope or wood to swell. A remarkable example of the force exerted by the swelling of lumber when wet is that illustrated in Fig. 100.

111. **Explanation of Capillary Action.** The forces that give rise to capillary action are those of *adhesion, cohesion,* and *surface tension.* In order to get at the reason for the elevation or depression of a liquid in a capillary tube let us consider a capillary tube highly magnified, as shown in sectional outline in Fig. 101.

If the adhesion between the tube and the liquid is greater than the cohesion between the molecules of the liquid, the liquid will wet the tube and will be drawn up around the edge, producing a concave surface. By concave surface we mean one that is depressed downward, Fig. 101. Now when such a concave surface is produced, surface tension comes into play, giving rise to an elevation of the liquid. We have learned that the surface of every liquid is under a state of tension, acting like a stretched membrane. Surface tension always tends to force the free surface of a liquid into the smallest area, which in this case is a flat surface, represented by the dotted line, *ab*. The force of surface tension then pulls the molecules composing the upper layer of the liquid toward the line *ab* and these in turn drag the molecules below them along by cohesion. But when the concave surface begins to move upward and flatten out, adhesion

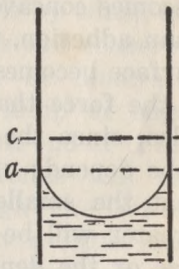


FIG. 101

again elevates the edges, producing a new concave surface; and so the liquid continues to rise from *ab* to *cd* and so on, until the force of surface tension pulling upward is just counter-balanced by the weight of the liquid in the tube.

Now just the reverse of this process takes place in a tube thrust into mercury. In mercury the cohesive forces acting between the molecules are relatively great as compared to the adhesive force between glass and mercury. When cohesion is greater than adhesion, the liquid does not wet the tube, and the resulting surface is convex. In this case surface tension tends to flatten the curved surface downward thus producing a depression of the liquid in the tube. The liquid is forced downward until its surface tension is just counterbalanced by the hydrostatic pressure of the liquid outside of the tube.

Thus it appears first, that, when adhesion is greater than cohesion, the liquid wets the tube and the surface becomes concave; second, that, when cohesion is greater than adhesion, the liquid does not wet the tube and the surface becomes convex; and third, that surface tension is the force that pulls the liquid up or pushes it down. Also, since the height to which a liquid will rise in a tube depends upon the weight of the column, it follows that the smaller the tube in cross sectional area, the greater will be the height to which a given liquid will rise or the depth to which it will be depressed. These facts may be stated concisely in the following laws.

112. Laws of Capillary Action. I. *Liquids rise in tubes which they wet, and are depressed in tubes which they do not wet.*

II. *The elevation or depression of the liquid is inversely proportional to the diameter of the tube.*

III. *The ascension or depression tends to decrease as the temperature of the liquid increases.*

113. Capillary Action in Soils. The distribution of moisture in the soil depends to a large extent upon capillary action. When the soil is compact, the minute spaces between the soil particles act as capillary tubes, thus aiding the water to rise to the surface. As the water evaporates from the surface, more of it rises by capillary action from the damper soil below, until the ground is thoroughly dried out.

If, on the other hand, the top soil is kept loose by cultivation, the spaces between the particles are too large for much capillary action to take place, and thus the moisture in the soil is conserved. In the semi-arid regions of the West, "dry farming" is successfully practiced. This consists in keeping the surface covered with a "dust mulch" produced by frequent cultivation. In this way the moisture is kept below the surface, where it can be utilized during the hot dry summer by the roots of growing plants.



FIG. 102.—Soil in footprint is dry and hard, due to capillary action

The principle by which cultivation of the soil conserves moisture may be illustrated by taking a lump of sugar and dipping one end of it into coffee. The coffee will rise rapidly through the whole lump. If now a little loose sugar be sprinkled over the top of the lump, with a spoon, it will be found that the coffee when it rises to this point, will take considerable time to pass through this loose sugar. This is the principle of the mulch in saving moisture. In Fig. 102 there is shown the impression of a shoe in a cultivated field. Where the earth is packed under foot

capillary action has taken place and the soil is dry and hard; around the footprint the soil is loose and underneath it is moist.

EXERCISES. **7.** A plant in a porous flower pot may be watered by placing the pot in a shallow dish of water. Explain.

8. Explain how soil water gets to the surface.

9. Why does soil that is packed at the surface lose moisture more readily than soil that is loose?

114. Osmosis. If into a thistle tube, over the end of which there has been fastened a membrane of a semipermeable nature, there be put a solution of sugar and water and the tube immersed in a beaker of water, Fig. 103, it will be observed that after a time the liquid in the tube will begin to rise. The same effect may be observed if, instead of sugar, a solution of almost any salt in water is placed in the tube. The water in the beaker passes through the membrane more readily than the solution (sugar or salt) passes out. The rise of the liquid in the tube is due to a pressure called *osmotic pressure*. The process by which liquids pass through membranes is called *osmosis*. Osmosis is due to molecular forces and depends upon the character of the liquids separated by the membrane, and also upon the nature of the membrane itself.

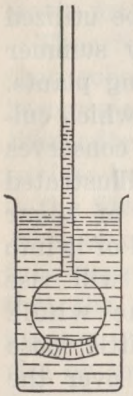


FIG. 103
Osmotic pressure

The principle of osmosis plays a most important part in the life of plants and animals. The roots of plants contain liquids which have a greater concentration than soil water. The thin-walled surfaces of the rootlets act as membranes through which the water in the soil passes. Osmotic action also takes place through the membranes in the bodies of human beings and other animals.

Molecular Forces in Solids

115. Characteristic Properties of Solids. The molecules of a solid are packed very closely together; their motions are therefore very much retarded. On the other hand, the forces which act between the molecules of a solid are enormously greater than those which act between the molecules of a fluid. Fluids of different kinds and densities diffuse readily into one another when brought into contact. Solids possess this property of diffusion to a very limited degree. The only case of the diffusion of solids at ordinary temperatures is that which has been observed to take place between gold and lead. If a flat piece of lead be placed upon a simlar piece of gold so that their surfaces come into intimate contact, the heavy molecules of gold will in time diffuse upward into the lead. These two metals are the only ones which exhibit this property when cold. At higher temperatures, however, all metals, as might be expected, show this property to a marked degree.

The physical properties of a metal are affected by the treatment to which it is subjected and the temperature, which it undergoes, changes. For example, the tenacity of iron is increased by drawing it into wire, and a steel spring under continued use will become more crystalline in structure and hence more brittle.

116. Adsorption of Liquids and Gases by Solids. Adsorption is the condensation of a fluid in a closely adhering layer upon the surface of a solid. Porous solids adsorb liquids and gases. The adsorption of a liquid by a solid may be illustrated by means of a device shown in Fig. 104. The apparatus consists of a porous cup into the end of which is fitted, by means of a rubber stopper, a bent glass tube containing a colored liquid. If the cup be placed in a beaker of water, the colored liquid will rise

in the outer arm of the tube, thus showing an increased pressure inside the cup due to the adsorption of water by its walls.

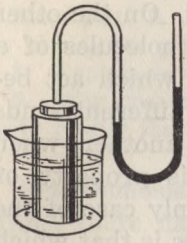


FIG. 104
Adsorption of a liquid by a solid

Some solids have the power to adsorb gases to a remarkable degree. This is especially true of freshly burned charcoal, as may be demonstrated by means of an apparatus as shown in Fig. 105. Some freshly burned charcoal, cooled to room temperature, is put into a wide-mouthed bottle filled with carbon dioxide, CO_2. The bottle is then closed with a rubber stopper into which is fitted a bent glass tube as shown. The lower end of the tube dips into a beaker of colored liquid. The charcoal quickly adsorbs a considerable portion of the carbon dioxide, thus creating a partial vacuum which is manifested by a rapid rise of the colored liquid in the tube.

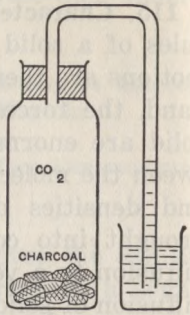


FIG. 105
Adsorption of carbon dioxide by charcoal

117. Crystallization. When matter changes from the liquid to the solid state, it may assume one of two forms so far as its molecular structure is concerned; it may be either *crystalline* in its structure or *amorphous*. When the molecules are arranged according to a definite form or pattern, the substance is said to be in a crystalline condition. When the molecules have no definite arrangement, the substance is said to be amorphous. Ice is a good example of a crystalline substance. The formation of frost on the window panes in winter, the feathery crystalline forms seen in snow flakes, and the beautiful crystalline structures which hoarfrost sometimes as-

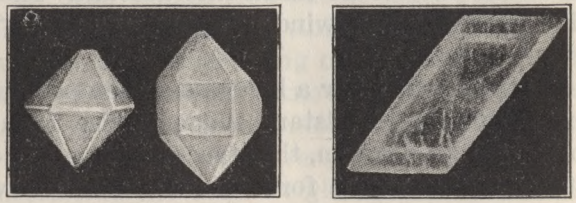
sumes, Fig. 106, gives some idea of the almost infinite variety of crystalline forms which water may assume on freezing. Many crystalline substances assume definite geometrical forms, as illustrated by crystals of quartz, Fig. 107 *I*, and gypsum, *II*. The nature of a substance is often determined by the crystalline forms which it assumes.



FIG. 106.—Crystals of hoarfrost on a wire fence

Blackboard crayon is an example of an amorphous substance. Most forms of glass, such as that used in test tubes, beakers and window glass are amorphous.

Some substances may occur in both the crystalline and amorphous conditions. This is especially true of carbon and iron. Diamond, for example, is pure carbon in a crystalline condition, while



I *II*

FIG. 107.—Crystals of quartz (*I*) and gypsum (*II*)

charcoal, which is incompletely burned organic matter, is carbon in the amorphous form. Cast iron is crystalline, while wrought iron is amorphous.

In general, when substances crystallize they expand. Water, for example, in freezing exerts an enormous ex-

pansive force, as is seen in the bursting of water pipes in winter, the upheaval of cement walks, and the disintegration of rocks. Indeed, the major portion of the soil is formed by the crumbling of the rocky material of the earth's crust, due in a large part to the expansive force exerted by freezing water.

Strength and Elasticity of Solids

118. Introductory. When an engineer undertakes to design a machine, a bridge, or a building, it is necessary for him to take into account, among others, three things. These are (a) the load which the proposed machine or structure is designed to carry, (b) the elasticity of the various members (shafts, girders, or beams), and (c) the breaking or tensile strength of the materials used.

In the case of a machine the *load* is the resistance which is to be overcome, and in the case of bridges and buildings and similar structures it is the weight of the structure itself plus the weight of its contents. In addition to the "dead weight" load, that is, the weight of the structure and its contents at rest, there are other force factors which have to be taken into account. These are the "live weight" load, wind pressure, and the "factor of safety."

When a train runs over a bridge or when people in an auditorium or in a grandstand suddenly rise from a sitting to a standing position, the structures in question are subjected not only to the force due to the actual weight of the train or the people, but they have to withstand also the forces generated by the sudden and somewhat violent motions of these bodies. A "live weight" load is a force equal to the weight of the structure and the weight of the body (train or people) plus the force due to the motion of the body. The distinction between dead and live weight loads may be demonstrated by means of

a platform scale such as shown in Fig. 108. The man on the scale notes his weight while in a crouching position. He then suddenly rises to a standing position, whereupon the scale shows an apparent increase of weight. The force recorded when he is in a crouching position and at rest is the dead weight; that recorded while he is in the act of rising is the live weight. It has been shown by experiment that the act of suddenly standing up from a crouching position on a platform scale may add as much as 60 per cent to the force due to the dead weight load.

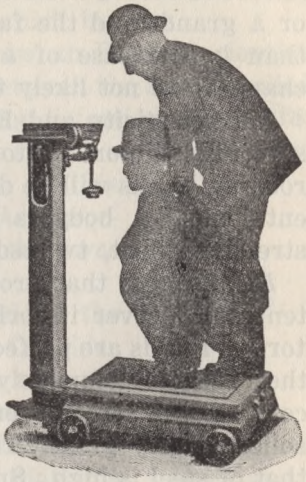


FIG. 108.—The act of suddenly rising from a crouching position may add over 60 per cent to one's "dead" weight

In designing a structure the engineer must not only make all parts strong enough to stand the dead weight load but he must provide also for such additional forces to which it may be subjected. This additional strengthening of the various members of a structure involve what is known as the factor of safety. The *factor of safety* is the ratio of the load actually provided for (including wind pressure, overloading, and live load effects) to the dead weight load; that is,

$$\text{factor of safety} = \frac{\text{load provided for}}{\text{dead weight load}}$$

For example, a brick wall is constructed to support a dead weight load of 15 tons. To provide for possible overloading, the wall is made strong enough to support a load of 150 tons. The factor of safety in this case is $150/15 = 10$. In the use of steel, the factor of safety is

usually 4. This factor varies with the material and the nature of the load. In the case of a bridge, an auditorium, or a grandstand the factor of safety is usually greater than in the case of a building in which violent load changes are not likely to occur.

119. Elasticity and Elastic Limit. In mechanical design it is important to know how much cables, beams, rods, or blocks will be distorted when subjected to different loads. A body is said to be distorted when it is stretched, bent, twisted, or compressed.

Elasticity is that property by virtue of which a body tends to recover its original conditions after being distorted. Fluids are perfectly elastic; that is, if compressed, they recover completely their original volume when the compressing force is removed. The range of elasticity in solids is very great. The elasticity of lead is low while that of steel is high. Springs are usually made of highly tempered steel because of its elastic properties. One of the properties of rubber, which makes it so valuable in the many uses to which it is put, is its remarkable elasticity.

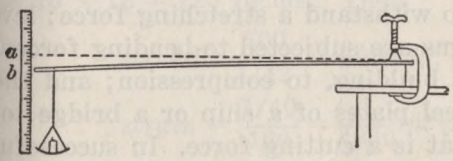
Limit of elasticity. Let a rod be clamped firmly at one end, and a force sufficiently great applied to the other end so that it is distorted (bent, stretched, or twisted). Now remove the distorting force. If the rod returns to its original condition with respect to its shape or volume, we say that it was distorted *within* its elastic limit. If, on the other hand, the rod is distorted to such a degree that it does not return to its original condition, that is, it has acquired a "permanent set," we say that it was distorted *beyond* its elastic limit.

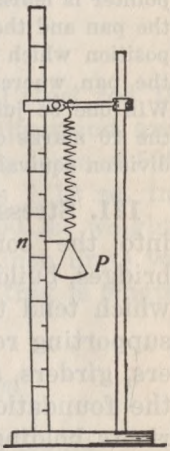
Different substances differ very widely in the amount of distortion which they can undergo before reaching the elastic limit. In the case of lead, the elastic limit is very small; in the case of hard drawn brass or tempered

steel, it is relatively great. It is obviously very important for the engineer to have exact knowledge of the elastic limits of the various materials which he uses.

120. Hooke's Law. Robert Hooke, an English scientist, in 1676 formulated the relation between the force applied to an elastic body and the resulting distortion. His statement of this relation later became known as Hooke's Law, which may be stated as follows: *Within the elastic limits of a substance, the distortion produced is proportional to the force applied.*

Hooke's law may be illustrated by the apparatus of Fig. 109, *I* and *II*. Consider first *I*. A rod is clamped at one end, the other being free to bend. The zero position of the free end is noted, and then a given weight is added. Suppose that this





I *II*

FIG. 109.—Apparatus for illustrating Hooke's Law

causes a bending of 1 cm. Now if twice as much weight be added, the rod will bend 2 cm, and so on until the elastic limit is reached. According to Hooke's law then, if a given force will bend, twist, or stretch a body through unit distortion, the application of twice as much force will produce twice the bending, twisting, or stretching; three times the force, three times the distortion, and so on until the limit of elasticity is reached. The apparatus, shown in Fig. 109 *II*, enables one to verify Hooke's law by stretching.

EXERCISES. **10.** It is found that a force of 10 lb. will stretch a given wire 0.01 in. What stretch will be produced by a force of 50 lb., provided the distortion is within the elastic limits of the substance?

11. Assuming that Hooke's Law holds, what force will be required to stretch the wire (Ex. 10) 0.1 in.?

12. It is desired to determine a set of scale readings for a spring balance. The spring (Fig. 109, *II*) with pointer and pan attached is fixed in a vertical position. The resting position of the pointer is marked 0 on the scale. A 10 g weight is then placed in the pan and the pointer moves down, coming to rest in the second position which is marked 10. The weight is now removed from the pan whereupon the pointer goes back to the zero position. Will one be justified in dividing the distance between the 0 and the 10 marks on the scale into ten equal divisions, calling each division equivalent to 1 g? Explain.

121. Stress and Strain. The materials which enter into the construction of machines, ships, airplanes, bridges, buildings, etc. are constantly subjected to forces which tend to produce distortions. Thus a cable or a supporting rod has to withstand a stretching force; levers, girders, and beams are subjected to bending forces; the foundations of a building, to compression; and the rivets holding the steel plates of a ship or a bridge together to a shear, that is a cutting force. In successful mechanical designing it is absolutely necessary that the designer take into account not only the kind, size, and shape of the various members of the proposed structure, but also that he select materials which are best suited to resist the forces applied. For example, a brick or a concrete wall is well suited to withstand compression but is not at all suited to resist bending forces; cast iron is about four times as well suited to withstand compression as it is to resist a stretching force; a steel wire cable is ideal for tension (stretching force) but is useless to resist a bending force, and so on. In other words the engineer has to keep constantly in mind the stresses to which the

various parts of the proposed structure are to be subjected and the resulting strains. The term *stress* refers to the force applied and *strain* to the effect (distortion) produced. For example, in the case of a rod subjected to a stretching force, stress and strain may be defined as follows:

$$stress = \frac{force\ applied}{cross\ sectional\ area\ of\ rod}$$

$$strain = \frac{stretch}{length\ of\ rod}$$

Example. Suppose that we have a vertical steel rod 10 feet long which is fastened securely at the upper end, and to the lower end there is applied a stretching force of 500 lb. The cross sectional area of the rod is 1/12 sq. in. It is stretched 1/40 of an inch by the 500 lb. weight. Since the cross sectional area is given in inch units we must express the length in the same units, that is length $= 10 \times 12 = 120$ inches. Then

$$stress = \frac{500}{1/12} = 6000\ lb.\ per\ sq.\ in.$$

$$strain = \frac{1/40}{120} = \frac{1}{4800} in.\ per\ in.\ of\ length$$

122. Breaking or Tensile Strength. The *tensile strength* of a substance is the force required to pull apart a bar of the material having unit cross section area. For example, when we say that the tensile strength of a given sample of steel is 60,000 pounds per square inch we mean that a force of 60,000 pounds is required to pull apart a bar of this metal having a cross section of one square inch.

The tensile strength of a substance is affected by its composition, and by the treatment to which it has been subjected. A slight change in the carbon content of steel may affect its tensile strength enormously. Thus a spec-

imen of mild steel containing 0.2 per cent of carbon has a tensile strength of 70,000 pounds per square inch, while another specimen having a carbon content of 0.8 per cent has a tensile strength of 140,000 pounds per square inch. The tensile strength is also affected by annealing, by tempering, and by drawing into fine wire.

Some especially prepared steels, such as those which are used in the manufacture of piano wires, have a tensile strength as high as 300,000 pounds per square inch.

123. Stiffness and Strength of Beams. In the use of rods and beams in building operations it is often desirable to secure the greatest strength and stiffness with the least weight of material. To accomplish this, beams are not made of compact form, but rather of a cross section like the patterns shown in Fig. 110, since a given amount of material is more effective in resisting a bend or a twist when placed some distance from its center of cross section. For this reason the steel "I" beam is commonly used in the construction of buildings. When it is necessary to combine lightness with great stiffness, the tubular form may be used, as in the frame of the bicycle. In nature we find many examples of this combination of strength and lightness, as in the case of the hollow stalks of grain, the bones of the body, and the quills of birds.

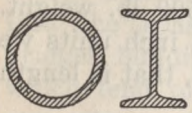


FIG. 110
Stiffness and strength of beams

It must not be understood, however, that a hollow beam is stronger than a solid one of the same dimensions. A hollow beam is stronger than a solid one of the same *weight,* but not of the same dimensions. If the saving of weight and material were no object, beams would always be made solid.

SUMMARY

1. Cohesion is attraction between like molecules. Adhesion is attraction between unlike molecules.

2. Viscosity is that property of matter in a fluid or semi-fluid state which tends to retard its property of flowing. Viscosity is due to the friction which is encountered by the molecules as they roll or slide one over the other.

3. Malleability is that property by virtue of which a substance may be beaten out into a thin sheet.

4. Ductility is that property by virtue of which a substance may be drawn out into thread-like strands.

5. Tenacity is that property of a substance by virtue of which it resists being torn apart.

6. Elasticity is that property of a substance by virtue of which it tends to recover its shape or volume after being stretched, bent, twisted, or compressed.

7. Kinetic Theory of Gases: According to this theory the molecules of a gas are in constant motion, striking against each other and against the sides of the containing vessel. The speed of the molecules of the gas depends upon the mass of the molecules and the temperature. Pressure exerted by a gas is due to forces exerted by the moving molecules.

8. Surface Tension: The surface of every liquid is under a state of tension, and acts exactly as if it were a stretched membrane. Surface tension is a force which acts parallel to the surface of the liquid.

9. Capillary action is that phenomenon which is associated with the elevation or depression of liquids in capillary tubes. If a liquid wets the tube, the surface of the liquid in the tube is concave and the liquid rises; if the liquid does not wet the tube, the surface is convex and the liquid is depressed.

10. Capillary action depends upon three factors: cohesion, adhesion, and surface tension. The force which causes elevation or depression is due to surface tension.

11. Laws of Capillary Action: I. Liquids rise in tubes which they wet, and are depressed in tubes which they do not wet. II. The elevation or depression of a liquid in a capillary tube is inversely proportional to the diameter of the tube. III. The ascension or depression tends to decrease as the temperature of the liquid increases.

12. Osmotic Pressure and Osmosis: Osmotic pressure is due to the unequal passage of liquids through a membrane. The process by which a liquid passes through a membrane is called osmosis.

13. Crystalline and Amorphous Conditions of Matter: When

the molecules of a substance in the solid state take on a definite form or pattern the substance is said to be crystalline. When the molecules do not take on any definite form or pattern, the substance is said to be amorphous.

14. Elasticity and Elastic Limits: Elasticity is that property of a substance by virtue of which it tends to recover its original condition when it is distorted. A body is distorted when it is stretched, bent, twisted, or compressed. When a body fails to recover its original condition, it is said to be distorted beyond its elastic limits.

15. Hooke's Law: Within the elastic limits of a substance, the distortion produced is proportional to the force applied.

16. Stress and Strain: Stress refers to the force applied and strain to the distortion produced. Expressed in equational form

$$stress = \frac{force\ applied}{cross\ sectional\ area}$$

$$strain = \frac{distortion}{length\ or\ volume}$$

17. Tensile Strength: The tensile strength of a substance is the force required to pull apart a bar of a given material having a unit cross sectional area. In the English system tensile strength is rated in pounds per square inch, and in the metric system in grams or kilograms per square centimeter.

REVIEW EXERCISES

A

1. What is a general property of matter? a special property? Give examples of each.

2. What is cohesion? adhesion? Why cannot a piece of crockery ware that is broken be mended by pressing the parts together?

3. Why does a blacksmith heat pieces of iron which he wishes to unite red hot before welding them?

4. What is the cause of viscosity in liquids? Why is the viscosity of gases less than that of liquids?

5. Define malleability, and name some substances that possess this property to a high degree.

6. What is ductility? Name some very ductile metals.

7. What is tenacity? Name two metals that have relatively high tenacities; one that has a very low tenacity.

8. Define elasticity. How does the elasticity of solids compare with that of fluids?

9. Discuss briefly the kinetic theory of gases, and describe an experiment upon which this theory is based.

10. Describe an experiment designed to illustrate the tendency of gases to expand.

11. How do you account for the fact that the gases of which the atmosphere is composed do not expand indefinitely into space?

12. Explain why a gas confined in a vessel exerts pressure upon the sides of the vessel.

13. Why does decreasing the volume of a gas by compression increase the pressure which it exerts?

14. Describe an experiment illustrating the diffusion of gases through solids.

15. Define surface tension, and describe an experiment which demonstrates the fact that the surface of a liquid tends to contract.

16. Explain why the drops of a liquid always tend to assume a spherical form.

17. What happens when the end of a capillary tube is thrust into (a) water? (b) mercury?

18. Give five examples of capillary action. If possible, give examples other than those mentioned in the text.

19. What is the relation of adhesion, cohesion, and surface tension to capillary action? What causes the surface of the liquid in the capillary tube to become curved? What is the force which causes the liquid to rise or fall in the tube?

20. State the laws of capillary action.

21. Explain what is meant by the elastic limit of a substance.

22. State Hooke's Law, and describe an experiment to illustrate this law.

23. In the calibration of a spring balance what assumption is made regarding the elasticity of the spring?

24. If a load of 10 g will cause the pointer of a spring balance to move downward through a distance of 35 mm, what displacement will a load of 4 g produce?

B

25. Explain the application of surface tension to the manufacture of lead shot.

26. Given two capillary tubes, *A* and *B*. The inside diameter

of *A* is 0.4 mm and that of *B* 0.2 mm. In which tube will water rise the higher, and how much?

27. Describe an experiment illustrating osmotic pressure.

28. What is the distinction between the crystalline and amorphous structures of matter? Name two crystalline substances and two amorphous substances.

29. What is the difference between "dead weight" and "live weight" loads, as applied to structures?

30. What is meant by tensile strength of a substance? How is the tensile strength of steel affected by its carbon content?

31. Two steel rods, *A* and *B*, have the same tensile strength. The diameter of *A* is 1 cm and that of *B* 2 cm. If a force of 20,000 lb. is required to break rod *A*, what force will be required to break rod *B*?

C

32. A load of 2000 lb. is supported by a vertical rod having a cross sectional area of 0.2 sq. in. Compute the stress.

33. A rod 20 ft. long is stretched 0.25 in. by a given force. Compute the strain.

NOTE: For additional problems, see Appendix.

STOP for TEST

START for TEST

CHAPTER V

MECHANICS OF SOLIDS

Motion, Velocity, and Acceleration

124. Rest and Motion Relative Terms. Whenever a force acts upon a body, it always tends to set the body in motion. A body is said to be in motion when it changes its position with respect to some fixed point. Rest and motion are relative terms; that is to say, a body may be at rest with reference to one object and in motion with reference to another. For example, a person sitting quietly in his seat in a moving car is at rest with respect to the seat and in motion with respect to objects outside.

Exercises. **1.** If the hand be clenched and moved about, are the fingers at rest or in motion (a) with respect to each other? (b) with respect to the body?

2. Consider the wheel of an automobile which is in motion. Are the spokes of the wheel in motion or at rest with respect (a) to each other? (b) to the ground?

3. Give another example of relative motion and rest.

125. Kinds of Motion. Motion may be classified in two ways, first, with reference to its *direction,* and second, with reference to its *rate.* With reference to direction

$$\textit{motion may be} \begin{cases} \textit{straight} = \textit{rectilinear} \\ \textit{curved} = \textit{curvilinear} \end{cases}.$$

With reference to its rate of change

$$\textit{motion may be} \begin{cases} \textit{uniform} \\ \textit{accelerated} \end{cases}$$

When a body travels over equal spaces in equal intervals of time, its motion is uniform; when it travels over

unequal spaces in equal intervals of time, its motion is accelerated. There is probably no such thing in the universe as an example of absolutely uniform motion, since all bodies in moving change in some degree their rate. We can, however, imagine such a thing as uniform motion, or we may find examples of motions which are approximately uniform.

EXERCISE. **4.** Classify the following motions, first, as to rectilinear or curvilinear, and second, as to uniform or accelerated: (a) A train running at the rate of 50 mi. per hour on a straight, level track; (b) a ball rolling down an inclined plane; (c) the swinging pendulum bob of a clock; (d) the rotation of the earth on its axis.

126. Speed and Velocity. *Speed* is the rate of motion per unit of time, *without reference to direction.* When we speak of the speed of a race horse as it runs around the track, the speed of a bird in flight, or that of an automobile, the direction of motion in all three cases may be changing in some degree at every instant. Speed is a general term. It refers only to the rate of motion of a body without any special reference to its direction of motion.

Velocity, on the other hand, is rate of motion per unit of time, with reference to a definite direction. For example, we may speak of a body as having at a given instant a velocity of 64 feet per second vertically downward, or of another body as having a velocity of 120 feet per second in a horizontal direction. When speaking of velocities, we do not always mention the direction; but direction of motion is always implied. Thus, when we speak of the velocity of a falling body, we may not state that the direction is vertically downward; but the direction of motion is implied.

Two bodies may have equal speeds but different velocities, so far as direction of motion is concerned. Suppose,

for example, that two bodies, A and B, Fig. 111, are connected by means of a flexible cord which passes over a pulley. B, being heavier than A, moves downward, while A moves upward. The speed of A at a given instant is the same as that of B. The velocity of A, on the other hand, is vertically upward; while that of B is oppositely directed, vertically downward.

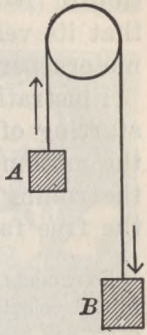


FIG. 111

Speed and velocity are usually expressed in miles per hour, feet per second, or centimeters per second, as the case may be.

Average speed or velocity is found by dividing the space passed over by the time; that is

$$\textit{average speed or velocity} = \frac{\textit{space}}{\textit{time}} = \frac{s}{t}$$

EXERCISES. **5.** The block C, Fig. 112, is drawn along the smooth surface of a table by means of the weight D, which is assumed to move downward with a uniform velocity of 30 cm per second. (a) How does the speed of C compare with that of D? (b) Compare the velocities of C and D with reference to magnitude (cm per second) and direction.

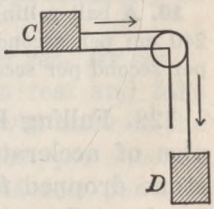


FIG. 112

6. A train runs 120 mi. in 4 hours. What is its average speed in (a) mi. per hour? (b) ft. per second?

127. Acceleration. When the velocity of a body increases or decreases, its motion is said to be accelerated. *Acceleration* is the change of velocity per unit of time. If there is an increase of velocity per unit of time, the acceleration is called positive; if a decrease, negative. When the unit of time is the second, velocity is expressed in units *per second;* acceleration, in units *per second per second.* Thus, when we say that a body has an accelera-

tion of 10 centimeters per second per second, we mean that its velocity has increased or decreased by 10 centimeters per second in one second.

Illustrations of accelerated motion are seen in the starting of a railway train or an automobile from rest, the motion of a sled or a toboggan down a hill or slide, the rolling of a marble down an inclined plane, and in the free fall of bodies.

EXERCISES. **7.** A sled starting from rest runs down hill. Its initial velocity is zero; its velocity at the end of the 1st second is 6 ft. per second; at the end of the 2d second, 12 ft. per second; at the end of the 3d second, 18 ft. per second, and so on. What is (a) the increase in velocity per second? (b) the positive acceleration?

8. A stone is projected vertically upward with an initial velocity of 4900 cm per second. Its velocity at the end of the 1st second is 3920 cm per second; at the end of the 2d second, 2940 cm per second, and so on. What is (a) the decrease in velocity per second? (b) the negative acceleration?

9. How many seconds will the stone continue to rise?

10. A ball rolling down an inclined plane acquires a velocity of 240 cm per second in 4 seconds. What is its acceleration in cm per second per second?

128. Falling Bodies. One of the most familiar examples of acclerated motion is that of falling bodies. A stone dropped from the hand falls to the earth with an accelerated motion due to the force of gravity. If it were not for the friction of the air all bodies would fall with equal velocities. If a feather and a coin, for example, be dropped from a given height, it will be observed that the coin will reach the ground first. If, however, the feather and coin be placed in a long tube and the air exhausted by means of an air pump, it will be found that both fall with the same velocity, showing that in a vacuum all bodies fall with equal velocities.

Galileo (1564-1642), an Italian physicist, was the first

to infer the truth of this statement. He had been taught that bodies fall with "velocities proportional to their weights." In order to test the truth or falsity of this statement, he dropped bodies of different sizes and densities from the top of the Leaning Tower of Pisa. He found that even very light bodies like paper fell with velocities which approached more and more nearly those of heavy bodies the more compactly they were wadded together. From these experiments he contended that all bodies, even the lightest, would fall at the same rate if it were not for the resistance of the air.



Galileo

129. Acceleration Due to Gravity. The acceleration imparted to a freely falling body by the force of gravity is represented by the symbol g. The numerical value of this acceleration due to gravity is approximately as follows:

$g = 32$ *ft. per sec. per sec.* $= 980$ *cm per sec. per sec.*

This means that if a body starts from rest and falls freely under the influence of gravity it will have at the end of the 1st second a velocity of 32 feet per second; at the end of the 2d second, $32 + 32 = 64$ feet per second; at the end of the 3d second, $3 \times 32 = 96$ feet per second, and so on. Or, if this be expressed in metric units, its velocity at the end of the 1st second will be 980 centimeters per second; at the end of the 2d second, $2 \times 980 = 1960$ centimeters per second, and so on.

The values of g given above are those which are usually used in making calculations involving the acceleration due to gravity. Since the force with which the earth attracts a unit of mass differs slightly from place to place, it follows that the value of g will be somewhat

different for different places. The student should bear in mind that the symbol g does not stand for gravity but for the *acceleration* due to gravity.

Values of g in cm per second per second

Boston, Mass.	980.38	Denver, Colo.	979.60
Chicago, Ill.	980.26	Pike's Peak, Colo.	978.94

For a more complete list of the values of g, see Table IX, Appendix.

130. A Study of Accelerated Motion. As an example of accelerated motion let us consider the case of body B falling vertically downward under the force of gravity, Fig. 113. Now in the study of accelerated motion there are five factors to be taken into account. These are: (a) the acceleration of the body; (b) the velocity which it has at the end of any given second; (c) the average velocity per second; (d) the space passed over per second; and (*e*) the total space passed over during any given period. In the outline given in Fig. 113 the entire period is 4 seconds. At first glance it would appear that the problem of acceleration is a very complicated one. The problem, however, becomes relatively simple, provided the student is willing to follow in detail the various steps involved. Referring to Fig. 113, let us consider these five factors in order.

(a) *Acceleration.* In order to avoid fractions we shall let $g = 32$ feet per second per second; that is, the body, starting from rest has a velocity gain of 32 feet each second of its fall.

(b) *Velocity.* Since the body starts from rest its velocity at the beginning is $V_0 = 0$. At the end of the 1st second its velocity $V_1 = 32$ feet per second; end of 2d second, $V_2 = 32 \times 2 = 64$ feet per second; end of 3d second, $V_3 = 32 \times 3 = 96$ feet per second; end of 4th second, $V_4 = 32 \times 4 = 128$ feet per second; and at the end of t seconds $V_t = 32 \times t = gt$ feet per second.

(c) *Average velocity.* Since the motion is uniformly accelerated, the *average velocity* for any given second or for any number of seconds may be found by dividing the velocity at the beginning of the period plus the velocity at the end by 2. Thus the average velocity during the first second $= \frac{1}{2}(0 + 32) = 16$ feet per second; 2d second, average velocity $= \frac{1}{2}(32 + 64) = 48$ feet per second; 3d second, $= \frac{1}{2}(64 + 96) = 80$ feet per second; 4th, $= \frac{1}{2}(96 + 128) = 112$ feet per second. In a like manner we may find the average for any number of seconds. For example, the average for the entire 4 seconds $= \frac{1}{2}(0 + 128) = 64$ feet per second. We may therefore write

$$\textit{average velocity} = \tfrac{1}{2}(\textit{vel. at beginning} + \textit{vel. at end})$$
$$= \tfrac{1}{2}(0 + gt) = \tfrac{1}{2}gt$$

(d) *Space passed over per second.* The distance traversed per second by the falling body will of course increase for each second of fall. The space passed over per second is equal to the average velocity times the time. But since the time in each case, as here considered, is 1 second, the number of feet traversed per second will be numerically equal to the average velocity for that second. Hence the space passed over during the 1st second $= 16$ feet; 2d second, $= 48$ feet; 3d, $= 80$ feet; 4th, $= 112$ feet, and so on.

(e) *Total space passed over during any given period.* There are two methods of finding the total number of feet passed over during any given number of seconds. Suppose that we wish to find the total space passed over during the 4 seconds. The first and most obvious method would be to get the sum of the distances traversed during the 1st, 2d, 3d, and 4th seconds of fall; that is, total space $S = 16 + 48 + 80 + 112 = 256$ feet. But since the space passed over in any given period is the average velocity for that period times the number of seconds, it

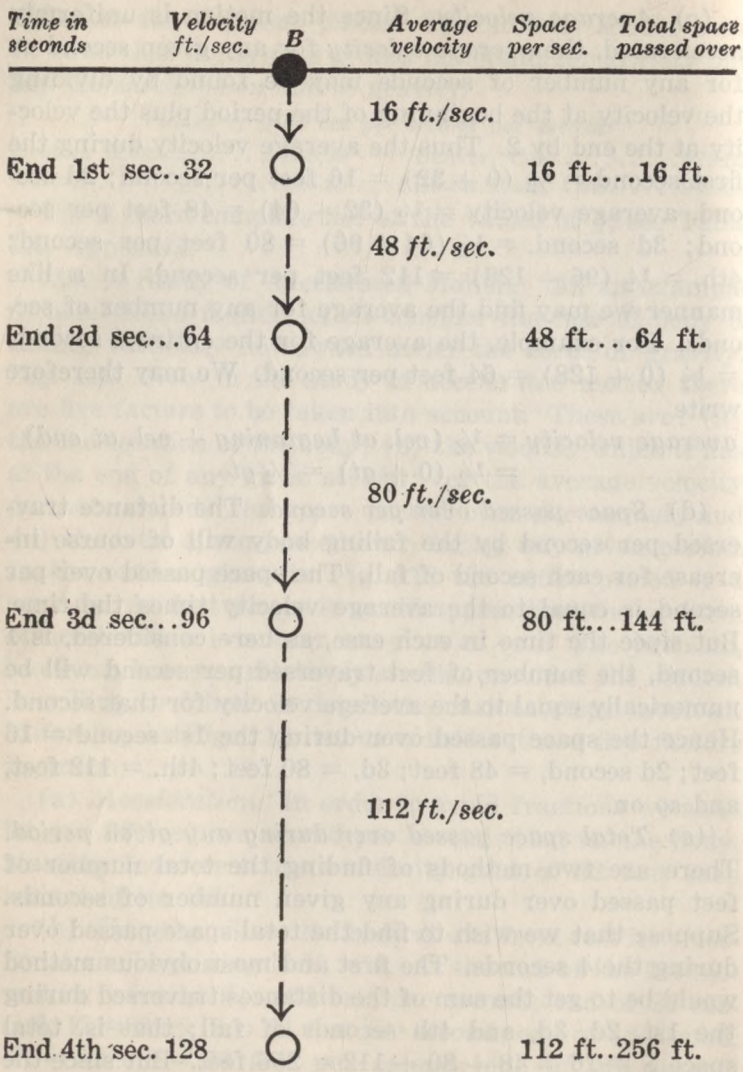


FIG. 113.—Velocities attained and spaces passed over by a body *B* falling from rest for 4 seconds

will at once occur to the student that instead of going to the trouble of finding the space passed over during each individual second and then getting the sum, a much more direct and convenient method would be to find the average velocity for the entire period and multiply the average velocity thus found by the number of seconds. Thus for the 4 seconds, the *average velocity* $= \frac{1}{2}(0 + 128) = 64$ *feet per second.* Hence it appears that the total space passed over is

$$s = \textit{average velocity} \times \textit{time}$$

But for a body starting from rest, the average velocity $= \frac{1}{2}gt$, therefore the total space passed over in t seconds is

$$s = \tfrac{1}{2}gt \times t = \tfrac{1}{2}gt^2$$

If now instead of using the acceleration of gravity g, as we have just done, we wish to consider the acceleration of any body in any direction, as for example, the acceleration of a ball rolling down an inclined plane or the acceleration of a car in starting or stopping, it is customary to use the letter a as a general symbol for acceleration. Our equations become

$$\textit{velocity} = \textit{acceleration} \times \textit{time}$$
$$v = at, \text{ and}$$
$$\textit{space} = \tfrac{1}{2}\ \textit{acceleration} \times \textit{time squared}$$
$$s = \tfrac{1}{2}at^2$$

131. Motion of Bodies Having an Initial Velocity. In our discussion of accelerated motion we have thus far considered only the case of bodies starting from rest, that is, bodies in which the initial velocity = 0. Suppose, however, that instead of starting from rest, a body is given an initial velocity v'. It is evident that in this case we shall have to take this initial velocity into account in writing our equations for accelerated motion. Thus at the end of any given interval of time the final velocity v attained by the body will be

$$v = v' + at$$

and the space passed over during the interval will be

$$s = v't + \tfrac{1}{2}at^2$$

If we take into account both positive and negative accelerations, we then have the general equations

$$v = v' \pm a\,t$$

$$s = v't \pm \tfrac{1}{2}at^2$$

where the + sign is used when the acceleration is positive and the — sign when the acceleration is negative.

Examples. 1. A body thrown downward from a very high building has imparted to it an initial velocity of 20 ft. per sec. Find (a) its velocity, and (b) the space passed over in 4 seconds, assuming the acceleration to be 32 ft. per sec. per sec. *Solution*: Since the acceleration is positive we have

(a) $v = 20 + (32 \times 4) = 148$ *ft. per sec.*, and

(b) $s = (20 \times 4) + (\tfrac{1}{2} \times 32 \times 16) = 336$ *ft.*

2. A body is projected vertically upward with an initial velocity of 250 ft. per sec. Find (a) its velocity at the end of the 4th second; (b) the space passed over during the 4 seconds. *Solution*: In this case the acceleration is negative; that is, there is a loss of velocity amounting to 32 ft. per sec. It follows then that

(a) $v = 250 - (32 \times 4) = 122$ *ft. per sec.*, and

(b) $s = (250 \times 4) - (\tfrac{1}{2} \times 32 \times 16) = 744$ *ft.*

EXERCISES. **11.** The data given in Fig. 113 may be tabulated as follows:

Time in seconds	*Velocity in feet per second*	*Total space passed over*
Beginning 1st second	0	0
End 1st second	32	16
End 2d second	64	64
End 3d second	96	144
End 4th second	128	256
End 5th second	?	?
End 10th second	?	?

Find the velocity and the total space passed over at the end of the 5th second.

12. Determine by the shortest method at your command (a) the velocity at the end of the 10th second, (b) the total space passed over.

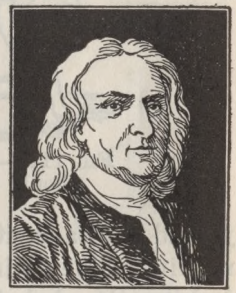
13. (a) What was the average velocity for the interval (3 seconds) included between the end of the 1st second and the end of the 4th? (b) Through what space did the body travel during these three seconds?

14. A ball rolling down an incline has an acceleration of 48 cm per sec/sec. (a) What is its velocity at the end of the 4th second? (b) How far will it travel during the 4 seconds?

15. An automobile starting from rest attained a speed of 12 mi. per hour in 8 seconds. Assuming that the gain in speed was uniform, what was (a) its acceleration in ft. per second? (b) its velocity in ft. per second at the end of the 6th second?

16. How far did the automobile (problem 15) travel during the 8 seconds?

132. Newton's Laws of Motion. Sir Isaac Newton (1642-1727), English mathematician and physicist, was one of the outstanding geniuses of his day. He constructed many new and unusual pieces of apparatus for use in his laboratory; he invented the method of the calculus in order to carry out certain mathematical computations in which he was engaged; he made important discoveries in Light, and formulated the law of universal gravitation. One of the most notable of his writings was on the Principles of Natural Philosophy, published in 1687 under the title *Principia.* In this famous document he proposed the three laws of motion which bear his name, and upon which the science of mechanics is largely based. These three fundamental physical laws state (a) the effect of force upon matter, (b) the relationship between force and acceleration, and (c) the equivalence of action and reaction.



Newton

Before taking up a detailed study of the three laws of motion, however, it will be necessary to consider three preliminary topics, namely, *inertia, momentum,* and *impulse.*

133. Inertia. *Inertia* is that property of matter by virtue of which a body at rest tends to remain at rest and when in motion tends to continue in motion in a straight line. In other words, inertia is the inability of a body to change its state of rest or motion. For example, a child is sitting on a sled. If now the sled is suddenly

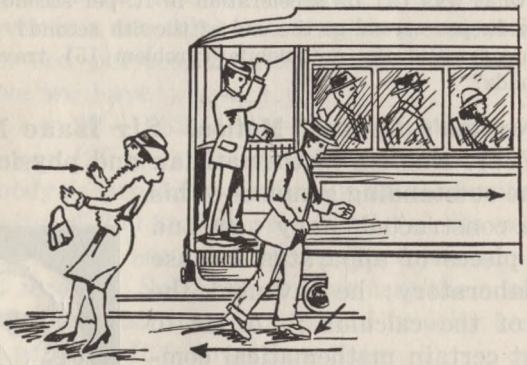


FIG. 114.—In stepping off a moving car one should face in the direction of motion

jerked forward, it is pulled from beneath the child. It requires time to impart motion to a body. If instead of jerking the sled, it had been put in motion slowly, this motion would have been imparted to the child and both would have moved forward. Not only does it take time to impart motion to a body, because of inertia, but it also takes time to stop a body which is in motion, as is illustrated by the forward motion of a person in a car when the latter comes to a sudden stop. In alighting from a moving car it is important to remember that inertia tends to carry the body forward in the direction

of the motion of the car. For this reason, one should always step off a car facing the direction of motion, as is shown in Fig. 114, and never facing to the rear.

Inertia may be illustrated by the card and ball experiment. Place a marble or small metal ball upon a stiff card, as shown in Fig. 115. When the spring is drawn back and then released the card is snapped out from under the ball, leaving it in position upon the vertical support.

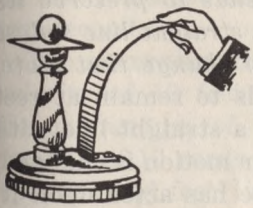


FIG. 115.—Inertia experiment

Other illustrations of inertia are seen in the stamping of snow from the feet; the beating of dust out of a carpet, or rather the beating of the carpet out of the dust; and in the "pounding" of a water pipe when the faucet is suddenly closed.

134. Hydraulic Ram. An illustration of the practical application of the principle of inertia is that furnished by the operation of the hydraulic ram. The *hydraulic ram* is an apparatus for elevating water to a height greater than its source. The condition necessary for the operation of the ram is that there be a supply of running water. The water flows through a pipe p and out through a valve v, Fig. 116. As the speed of the stream through the pipe increases, valve v suddenly closes, thus causing the water, because of the inertia, to force its way for an instant through the valve V into chamber C. The water in the pipe p now rebounds, relieving the pressure on the valve v, which falls of its own weight, thus opening the orifice at the end of the pipe. The process is now

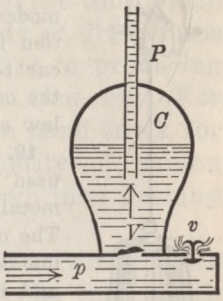


FIG. 116.—Hydraulic ram, a practical application of the principle of inertia

repeated, and more water is forced into chamber *C*. The air cushion in *C* tends to produce a steady flow in the vertical pipe *P*.

135. The First Law of Motion. We are now prepared to grasp the meaning of the first law of motion, which may be stated thus: *Every body tends to preserve its state of rest or uniform motion in a straight line unless compelled by some external force to change that state.* This means that a body at rest tends to remain at rest and if in motion it tends to move in a straight line with uniform velocity. If its state of rest or motion is changed, it is evidence that some outside force has acted upon it.

The first law of motion is sometimes called the law of inertia.

EXERCISES. **17.** An automobile is suddenly started with a jerk. (a) In what direction (forwards or backwards) are the occupants of the car thrown? (b) What part of the first law of motion is illustrated by the motion of the passengers?

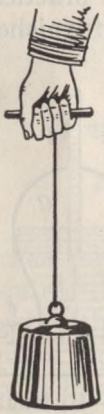


FIG. 117

18. The car (Ex. 17) is in motion, running at moderate speed, when the driver, seeing an obstruction in the road, applies the brakes, bringing the car to a sudden stop. (a) In what direction are the occupants thrown? (b) What part of the first law of motion is illustrated in this case?

19. One end of a piece of cotton twine, such as is used in tying bundles in stores, is fastened to a metal weight having a mass of 2 or 3 lb., Fig. 117. The other end of the string is wrapped around the middle of a lead pencil, or other convenient handle, and is firmly tied. By pulling gently upward on the string the weight may be lifted from the floor. If, however, a sudden pull be exerted, the string is broken. Explain.

136. Momentum. *Momentum* is the quantity of motion which a body possesses; it is equal to the mass of the moving body times its velocity. That is,

$$momentum = mass \times velocity = mv.$$

The unit of momentum is the momentum of unit mass having unit velocity. In the f. p. s. system the unit of momentum is the momentum of a mass of one pound having a velocity of one foot per second. In the c. g. s. system the unit of momentum is the momentum of a mass of one gram moving with a velocity of one centimeter per second.

The momentum of a mass of 10 pounds moving with a velocity of 10 feet per second is $10 \times 10 = 100$. Also, the momentum of a mass of 10 tons moving with a velocity of 10 yards per second, reduced to the same units (pounds and feet) is $2000 \times 10 \times 3 \times 10 = 600{,}000$.

EXERCISES. **20.** A 6 ton truck is moving at the rate of 30 mi. per hour. What is its momentum in f. p. s. units?

21. A mass of 5 kilograms falls freely for 4 seconds. What is (a) its velocity at the end of the 4 sec.? (b) its momentum in c. g. s. units?

137. Impulse. In estimating the effect of a force upon a body, it is necessary to consider two factors, namely, the magnitude (amount) of the force and the time which it acts. A small force acting upon a body of given mass for a long time may have the same effect in producing motion as a great force acting for a short time only. For example, a child pushing steadily upon a hand truck for 10 minutes may impart as much quantity of motion (momentum) to the truck as a full grown man pushing for only a minute or two.

We have a special name for this factor which takes into account both the force applied and the time which it acts; it is called impulse. *Impulse* is the product of force and time; that is,

$$impulse = force \times time = Ft$$

138. The Second Law of Motion. *Change of motion is proportional to the impressed force, and takes place in the direction in which the force acts.* This is the way in

which Newton stated his second law of motion. The meaning of the above statement, however, is not altogether clear without certain interpretations of the terms used. The two terms that are likely to cause trouble are "change of motion" and "impressed force." Maxwell, a noted English physicist, interpreted "change of motion" to mean *change of momentum* and "impressed force" to mean *impulse*. The first part of the law may then be written as follows:

$$\textit{impulse} = \textit{change of momentum}, \text{ that is}$$
$$\textit{force} \times \textit{time} = \textit{mass} \times \textit{change of velocity},$$

from which we have

$$\textit{force} = \frac{\textit{mass} \times \textit{change of velocity}}{\textit{time}}$$

But since change of velocity divided by time is acceleration, our equation becomes

$$\textit{force} = \textit{mass} \times \textit{acceleration}, \text{ or}$$
$$F = ma$$

This is an equation of great importance in physics. While its use and significance will be explained later on (Art. 146), it is important to note, first, that we have here a new way of defining force, not as a push or a pull, but as the product of mass and acceleration; and second, this definition of force comes directly from Newton's second law of motion.

The second law of motion is sometimes called the law of rate of change of momentum, and sometimes the law of acceleration. Both terms are very appropriate.

EXERCISES. **22.** A body of mass m is moving with a uniform velocity of 100 cm per sec. A force F now acts upon the body for 4 seconds, producing a final velocity of 320 cm per second. What was the momentum of the body (a) under the first condition? (b) at the end of the 4 seconds? (c) What was the total change of momentum?

23. In Ex. 22, what was (a) the change of velocity per second? (b) the acceleration in cm per sec. per sec.?

139. Further Discussion of the Second Law of Motion. The second law of motion not only tells us that change of motion is proportional to the impressed force but also that this change of motion is in the direction in which the force acts. Now this implies that if two or more forces act on a body at the same time, each force produces its own effect independent of the effects produced by the other forces. An illustration of this phase of the second law may be seen in the effect of gravity on a projectile, as shown in Fig. 118. Suppose that a ball is fired hori-

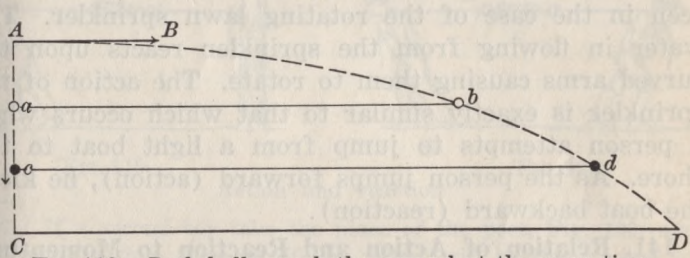


FIG. 118.—Both balls reach the ground at the same time

zontally from a gun in the direction AB. The ball is acted upon by two forces, that due to the explosion of the powder and that due to gravity. The force due to the powder acts horizontally in the direction from A toward B; the force of gravity acts vertically downward. The actual path of the ball is represented by the dotted line from B to D. Now suppose that at the instant the ball leaves the gun, a similar ball is dropped from A, falling vertically from A to C. Since a force has the same effect in producing motion whether the body upon which it acts is at rest or in motion, or whether the body is acted upon by that force alone or by other forces at the same time, it follows that both balls will reach the ground simultaneously. When one ball is at a, the other will be at b; and when the ball falling vertically has reached the position c, the second will be at d, and so on.

140. The Third Law of Motion. *To every action there is an equal and opposite reaction.* This is called the law of action and reaction. It tells us that every force is two-sided in its nature. If we push on a wall with a given force, the wall reacts (pushes back) with an equal force. In rowing a boat we act on the water, and the water in turn reacts on the boat, causing it to move. The bird in flying acts on the air with its wings; the air, on the other hand, reacts on the bird, giving it motion.

An excellent illustration of action and reaction is that seen in the case of the rotating lawn sprinkler. The water in flowing from the sprinkler reacts upon the curved arms causing them to rotate. The action of the sprinkler is exactly similar to that which occurs when a person attempts to jump from a light boat to the shore. As the person jumps forward (action), he kicks the boat backward (reaction).

141. Relation of Action and Reaction to Momentum. The terms action and reaction may be interpreted to mean *force*. When the powder in a gun explodes, it acts upon both the bullet and the gun with an equal force. Now force is proportional to change of momentum; and since the forces of action and reaction are equal, it follows that the momentum of the gun is equal to the momentum of the bullet. This relation of action to reaction may be expressed by an equation as follows:

$$mv = m'v'$$

in which m and v represent the mass and velocity of the gun and m' and v' the mass and velocity of the bullet. The velocity of the bullet will be as many times greater than the velocity of the gun as the mass of the gun is greater than the mass of the bullet. The lighter the gun, therefore, the greater its "kick" or recoil. If bullet and gun were of equal masses, they would fly apart when the powder explodes with equal velocities.

EXERCISES. **24.** Suppose that a ball of mass 1 lb. is fired from a cannon of mass 1 ton, with a velocity of 1000 ft. per second. What is the velocity of the gun's recoil, assuming that the momentum of the ball is equal to that of the gun?

25. When an apple falls to the earth, its momentum toward the earth, according to the third law of motion, is equal to the momentum of the earth toward the apple. How does the velocity of the earth compare with that of the apple?

26. A boy pulls on a rope, as shown in Fig. 119, with a force of 50 lb., as registered by the spring balance *S*. (a) What force is exerted on the rope? (b) What force is exerted on the post?

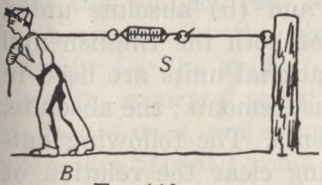


FIG. 119

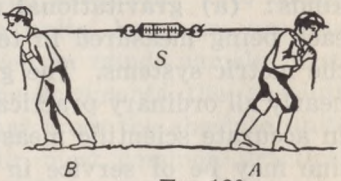


FIG. 120

Action and reaction

27. If a second boy take the place of the post, Fig. 120, and each pulls with a force of 50 lbs., what will be the reading of the spring balance?

142. Measurement of Force. The magnitude of a force may be measured in two ways (a) by the push or pull exerted, and (b) by the product of the mass moved and the acceleration imparted to it.

One of the most usual methods of measuring the pull resulting from the application of a force is by means of the spring balance, or dynamometer, as it is sometimes called. Attach to the hook of a spring balance a given mass and lift the mass against gravity. The pointer of the instrument moves to a certain position on the scale. This indicates, in pounds or grams as the case may be, the force with which the earth attracts the body. Now, attach the hook of the balance to some object, say a nail or a hook in the wall, and pull upon it. The reading of the instrument again indicates the magnitude of the force exerted.

The force exerted upon a given mass may also be measured by the acceleration which is imparted to it. This is the method usually employed in scientific determinations of the magnitude of a force. Thus to determine the magnitude of a force acting upon a body, it is necessary to know the mass of the body and its acceleration. The product of these two factors is a measure of the force; that is, $F = ma$.

The units of force employed in this country are of two kinds: (a) gravitational units and (b) absolute units, each being measured in terms of both the English and the metric systems. The gravitational units are used in nearly all ordinary practical measurements; the absolute, in accurate scientific measurements. The following outline may be of service in making clear the relation of the two sets of units.

Units of force
- *gravitational*
  - *pounds of force*
  - *grams of force*
- *absolute*
  - *poundals*
  - *dynes*

143. Gravitational Units of Force. The gravitational units of force are those which compare the push or pull exerted by a force with the attraction due to gravity for a given mass. The English gravitational unit is the force of a pound (also called the weight of a pound). The *force of a pound* is a force equivalent to the attraction of the earth for a pound mass. The *force of a gram* is a force equivalent to the attraction of the earth for a gram mass. Since the attraction of the earth for a given mass varies slightly for different places, the gravitational units of force likewise vary. The variation for ordinary practical measurements, however, is so small as to be negligible.

144. Absolute Units of Force. The absolute units of

force, the poundal and the dyne, are derived from the product of unit mass of a body and the acceleration imparted to it. The *poundal* is that force which will give to a pound mass an acceleration of one foot per second per second. The *dyne* is that force which will give to a gram mass an acceleration of one centimeter per second per second. Since the absolute units are used primarily in making accurate scientific measurements, it is not highly important, in an elementary work on physics, that this subject receive special attention. Two things with respect to the absolute units, however, are important for the student to keep in mind, namely, first, that in accurate scientific measurements the absolute units *are* used, and second, that from the absolute units there are derived some of our most common practical units, such as the *watt* and the *kilowatt*.

145. The Relation of Absolute to Gravitational Units. The relation of absolute to gravitational units may be expressed, approximately as follows:

1 *pound of force* = 32 *poundals*
1 *gram of force* = 980 *dynes*

The above values are arrived at thus. In gravitational units the force of a pound, for example, is the force with which the earth attracts a pound mass. Absolute units, on the other hand, are measured in terms of unit mass and unit acceleration; that is, $F = ma$. Now if a pound mass were allowed to fall freely in a vacuum, its acceleration at sea level would be about 32 feet per second per second (more accurately 32.16), and in a like manner, a mass of one gram would have an acceleration of about 980 centimeters per second per second. We say then that the force of a pound measured in gravitational units is equivalent to about 32 absolute units (poundals), and that a force of one gram in gravitational units is equivalent to about 980 absolute units (dynes).

EXERCISES. **28.** A mass of 2 pounds held in the hand exerts upon the hand, due to the attraction of the earth, a force of 2 pounds. What is the force in poundals?

29. A mass of 1 kilogram rests upon the table. What force does it exert upon the table in (a) grams of force? (b) dynes?

30. A magnet exerts a force of 6860 dynes on a piece of iron. Find the force in grams.

COMPOSITION AND RESOLUTION OF FORCES AND VELOCITIES

146. Introductory. A force is completely described when we know three things about it, (a) its point of application, (b) its magnitude, and (c) its direction. If, for instance, we say that a force of 20 pounds acts on a horizontal bar AB we have not fully described it, because we do not tell where the force is applied or in what direction it acts. If, on the other hand, we say that a force of 20 pounds is applied to the end A and in the direction from A to B, we have not only described the force fully but we have also given a cue as to the kind of motion which the force tends to impart to the body. In this case the bar tends to move horizontally in the direction of its length AB. In case the force had been applied vertically downward at the end A, the result would be a tendency to give the beam a rotary motion in a clock-wise sense.

Forces, velocities, and accelerations may be represented by straight lines drawn to scale.

147. Graphic Representation of Forces and Velocities. By way of illustration let us consider that a light body is floating upon the surface of a slowly moving stream of water. Suppose, too, that a wind is blowing in a direction at right angles to the direction of motion of the current. Assume that the force of the current carries the

floating body in the direction *AB*, Fig. 121, with a velocity of 80 feet per minute, and the force of the wind, blowing at right angles to the direction of the current, gives the body a velocity of 60 feet per minute in the direction *AC*. Now according to the second law of motion each force will have its effect in producing motion independent of the other. That is, in one minute's time the current will have carried the body down stream a distance of 80 feet, and in the same time the wind will have carried it in a direction at right angles to that of the current a distance of 60 feet. In order to satisfy these two conditions it is evident that the body will move along the path represented by the diagonal *AD*, arriving at *D* in one minute. We have here to consider the fact that the body is acted upon simultaneously by two forces which tend to give it two independent velocities, one in the direction *AB* and the other in the direction *AC*. The actual velocity is represented by the diagonal *AD*.

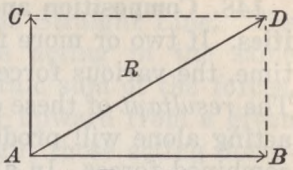


FIG. 121

In Fig. 121, the line *AB* represents the velocity due to the current, and the line *AC* (or its equivalent, *BD*) that due to the wind. If now we let a length of 1 inch represent a velocity of 10 feet per minute, then a line 8 inches long will represent the velocity due to the current, that is, a velocity of 80 feet per minute. And also, according to this scale, a line 6 inches long will represent a velocity of 60 feet per minute, the velocity due to the wind. Lines drawn to scale, then, may be used to represent velocities, accelerations, and forces with reference to magnitude and direction. When drawing upon the blackboard, the inch may conveniently be used as the scale unit; when drawing upon paper, the centimeter is the unit usually employed.

148. Composition and Resolution of Forces and Velocities. If two or more forces act upon a body at the same time, the various forces so acting are called *components*. The *resultant* of these components is a single force which acting alone will produce the same effect as that of the combined forces. In a like manner we may speak of the components and resultants of velocities and accelerations. For example, in Fig. 121, the lines AB and AC represent component velocities, and the line AD represents their resultant.

The *composition* of forces or velocities is the process of finding the resultant force or velocity. When the resultant is given, the process of finding the components is called *resolution*. Composition of forces or velocities means the combining of the forces or velocities into a single resultant. Resolution of forces or velocities means the separation of the resultant into its components.

The composition (process of finding the resultant) of forces or velocities may be accomplished either by graphic or by mathematical methods. In Fig. 121, for example, the components AB and AC are at right angles to each other, therefore the resultant AD is the hypotenuse of a right angled triangle. It follows then that

$$AD = \sqrt{AB^2 + AC^2} = \sqrt{80^2 + 60^2} = 100$$

EXERCISES. **31.** Two forces, represented respectively as AB and AC, act simultaneously on a body at A. The force $AB = 120$ lb., and $AC = 80$ lb. Draw to scale on paper or upon the blackboard lines to represent these forces, assuming that they act at right angles to each other. Complete the parallelogram and draw the resultant. By means of your scale determine the magnitude of the resultant R.

32. Assume that the forces AB and AC (Ex. 31) make an angle of 60° with each other. Draw lines to represent these forces graphically. It will be necessary to use a protractor (Art. 25) in order to determine the angles accurately. Determine by means of your scale the magnitude of the resultant R in this case.

149. Parallel Forces in the Same Straight Line. The resultant of two or more forces acting in the same straight line is equal to the algebraic sum of the forces. Components acting to the right or upward from a given point are said to have the plus sign; those acting downward or to the left, the minus sign. For example, a steamboat is running up the Hudson river against a strong head wind. The force exerted on the water by the engine is 1200 units; that exerted on the boat by the wind down stream is 200 units; and that due to the stream, 300 units. Here we have three parallel forces acting upon the boat, the force exerted by the engine up stream and hence positive, and the negative forces of wind and current directed down stream. The resultant is therefore

$$R = +1200 - 200 - 300 = +700.$$

What is true of parallel forces in the same straight line is true also of velocities.

EXERCISES. **33.** If a man on top of a freight car, which moves with a velocity of 30 ft. per second, runs in the direction in which the car is moving with a velocity of 10 ft. per second, what is the resultant (actual) velocity?

34. Suppose that the man (Ex. 33) runs in the opposite direction with a velocity of 10 ft. per second, what will be the resultant velocity?

35. Three forces, one due to the tide, one due to the wind, and the third the force exerted by the engine, act in the same straight line upon a boat. The tide acts to the westward with a force of 4 units; the wind acts in the same direction with a force of 12 units; the engine drives the boat to the eastward with a force of 50 units. Find the magnitude and direction of the resultant.

150. Parallel Forces Not in the Same Straight Line. The resultant of two forces acting in the same direction but not in the same straight line is equal to the sum of the forces, and the point of application of the resultant divides the distance between the forces into lengths in-

versely proportional to the forces. This may be illustrated by means of two spring balances, S and S', and a weight, Fig. 122. Let a given weight W (12 lb., for example) be suspended at such a point B that the force on the spring balance S is 8 pounds and the force on S' is 4 pounds. The resultant $W = 8 + 4 = 12$ lb.

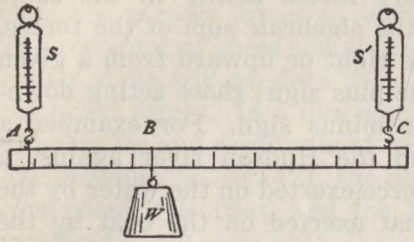


FIG. 122.—Parallel forces

It is important to note that the greater force 8 lb. is on the side of the short arm AB, and the lesser force 4 lb. is on the side of the long arm BC. The relation of the force acting on S to that acting on S' may be expressed as follows:

$$\frac{force\ on\ S}{force\ on\ S'} = \frac{BC}{AB}$$

Example. The supporting rod AC, Fig. 122, is 3 ft. in length. Where must a weight of 18 lb. be placed so that the force on S shall be 6 lb. and that on S' 12 lb.? *Solution*: Since the force on S is less than that on S' it follows that the long arm, AB, will be on the side of S, and the short arm, BC, on the side of S'. The length of the rod $AC = 3$ ft. $= 36$ in. Let $x =$ the long arm AB and $36 - x$ the short arm BC. Then

$$\frac{6}{12} = \frac{36 - x}{x}$$

Clearing of fractions, we have $6x = 12\ (36 - x)$, and hence

$$x = 24\ in. = AB$$
$$36 - x = 12\ in. = BC$$

EXERCISES. **36.** If W, Fig. 122, exert a force of 20 lb., what will

be the force on S and S' if W be placed (a) midway between A and C? (b) ¼ of the distance from A to C? (c) ⅕ of the distance from C to A?

37. If AC, Fig. 122, be 100 cm in length, where must a weight of 300 grams be placed so that the force exerted on S shall be 75 grams and that on S' 225 grams?

38. Two boys, Fig. 123, carry between them a block of ice weighing 90 lb. The distance AB is 2 ft. and the distance AC is 3 ft. Find what part of the load each boy carries.

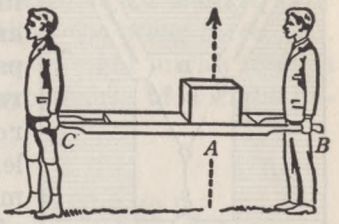


FIG. 123

151. Forces Making an Angle with Each Other. The resultant of two forces making an angle with each other is equal to the diagonal of the parallelogram formed by the lines representing the forces as sides. When the components act at right angles to each other, Fig. 121, the numerical value of the resultant R is equal to the square root of the sum of the squares of the other two sides.

If the angle formed by the components (forces or velocities) is not a right angle, the resultant R may be found by the graphic method. Suppose, for example, that A, Fig. 124, represents a boat on the surface of the ocean and that it is acted upon by two forces, the wind and the tide. The wind tends to give it a velocity in the direction AB of 20 feet per minute; the tide, acting at an angle of 60° to the direction of the wind, tends to give it a velocity of 15 feet per minute. Now assume that in this case we wish to find the direction and magnitude of the resultant *velocity*. In other words, we wish to find the actual direction and velocity of the boat. Lay off on paper a line

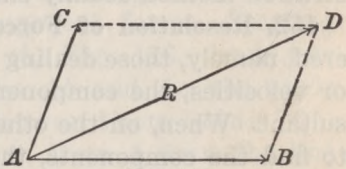


FIG. 124

AB, 20 centimeters in length, to represent a velocity of 20 feet. From the point *A*, Fig. 124, lay off *AC*, 15 centimeters in length, making with *AB* an angle of 60°. Now complete the parallelogram. The line *AD* represents the magnitude and direction of the resultant, its length in centimeters being numerically equal to the velocity of the boat in feet per minute.

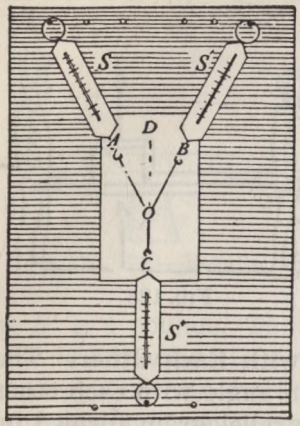


FIG. 125

When more than two components are given, the resultant may be determined by repetition of the parallelogram of forces.

The resultant of two forces may be determined experimentally by means of spring balances, as shown in Fig. 125. The resultant of the forces registered by balances *S* and *S'* is represented by the pull on balance *S''*. This resultant pull may be represented graphically by the line *OD*. This is the demonstration method usually employed in laboratory practice.

152. Resolution of Forces. In the cases just considered, namely, those dealing with the composition of forces or velocities, the components were given to find the resultant. When, on the other hand, the resultant is given to find the components, the process is called the *resolution* of forces or velocities, as the case may be. The problem of resolving a resultant force or velocity into its components in general requires the construction of a parallelogram having the given resultant as its diagonal. The components may be determined either by mathematical computation or by the graphic method. We shall take up for consideration two general cases.

The simplest case is that in which there is given the

resultant and one of the components to find the other, the two components acting at right angles to each other.

Example. The resultant R of two forces AB and AC acting at right angles to each other is 20 pounds. The component AB is 16 pounds. Find the other force, AC. Since the forces AB and AC are at right angles to each other, the resultant R is the hypotenuse of a right angled triangle. Therefore

$$AC = \sqrt{R^2 - AB^2}$$
$$= \sqrt{20^2 - 16^2} = \sqrt{144} = 12 \text{ } lb.$$

A more general case is that in which there is given the resultant R and the angle which it makes with one of the components, to find the magnitude and direction of the components. In elementary mechanics the easiest way to solve this problem is by the graphic method. Suppose that we wish to resolve a force R into two components, which make an angle of 70° with each other, under the condition that the resultant R makes an angle of 30° with the horizontal component. Given the resultant R equal to 120 pounds, say, to find the magnitude of the component forces. Starting at A, Fig. 126, draw two lines Ab and Ac, the angle between them being 70°. Next, from A draw the line AD to represent R, making an angle of 30° with the line Ab. If, now, we let 1 centimeter equal 10 pounds, then a line 12 centimeters will represent the magnitude of R, 120 pounds. Now complete the parallelogram by drawing the lines DB and DC. The lines AB and AC represent in magnitude and direction the components of the resultant R. The values of AB and AC may be determined from the scale chosen.

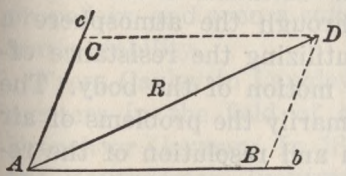


FIG. 126

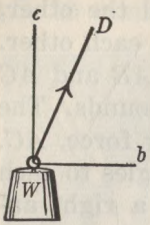


FIG. 127

EXERCISES. **39.** The weight W, Fig. 127, is acted upon by a force of 160 pounds which is directed from W to D, making an angle of 30° with the vertical line WC. Find by drawing to scale the horizontal and vertical components of the force WD.

40. A motor boat on a lake is moving from A to D in a north-easterly direction with a velocity of 12 mi. per hour. Find the northern and eastern velocity components of the boat.

41. The lake (Ex. 40) is 4 mi. wide from west to east. How long will it take the boat to reach the eastern shore?

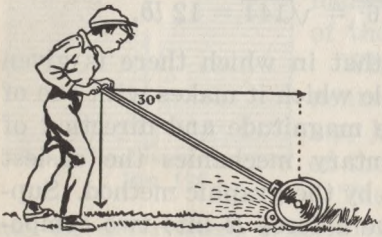


FIG. 128

42. A boy pushes a lawn mower with a force of 50 lb., the force acting in the direction of the handle, which makes an angle of 30° with the horizontal, Fig. 128. Find, by drawing, what part of the force acts, (a) horizontally, and (b) vertically.

THE AIRPLANE

153. The Story of Aviation. Aviation is the science and the art of propelling through the atmosphere a heavier-than-air machine by utilizing the resistance offered by the air itself to the motion of the body. The study of aviation involves primarily the problems of air reactions and the composition and resolution of the resulting forces.

For centuries man attempted to emulate the flight of birds, which are in a sense heavier-than-air machines. The soaring bird and the flying kite naturally were the forerunners of the modern airplane. Indeed the airplane may be regarded as a kite which is self propelling. The kite depends upon the wind for its support; the airplane, on the other hand, by means of its revolving propeller makes its own wind. The idea of flying is as old as

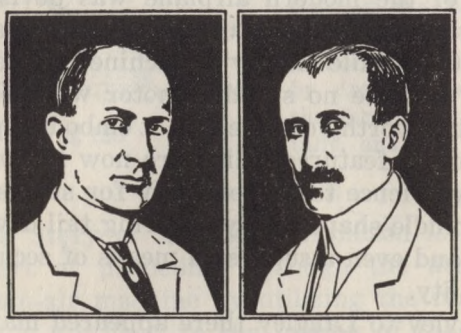
mythology. Our literature contains many accounts of man's attempts to "fly like a bird." Daedalus, mythical mechanical genius of the early Greeks, is reputed to have fabricated wings with which he flew from Crete to Sicily, and his son Icarus, who was flying with his father, forgetting the advice which had been given him, flew so high that the sun melted his wax wings and he fell into the sea and was drowned. Thus the fabled Icarus became the first victim upon the altar of aviation. And at a later date J. T. Trowbridge, in his famous poem, pictured the disastrous flight of Darius Green and his flying machine.

The first flying machine embodying the essential characteristics of the modern airplane was devised by Sir George Cayley, an Englishman, who published an account of it in 1809. While Cayley's machine did not actually fly, largely because no suitable motor was at that time available, it is worthy of note that it embodied practically all the essential features which are now known through practical experience to be necessary for successful flight —wings, spindle shaped body, steering tail device, screw propellers, and even a suggested means of securing automatic stability.

From Cayley to Langley there appeared many experimenters in the field of aviation, most of these being French or German. In 1896 Professor S. P. Langley of the Smithsonian Institution, Washington, D. C., succeeded in making a small steam driven model weighing about 30 pounds which accomplished one of the first successful airplane journeys, covering a distance of nearly a mile before running out of fuel. Langley's first model was not large enough to carry a pilot. He at once set about building a larger model, one which was designed to be operated by a pilot. In order to provide suitable motive power he designed and built a gasoline motor

which for that day was considered to be a marvel of lightness and power. He called his plane an "Aerodome," a name which has since been used to designate a building in which airplanes are stored. This machine was equipped with two sets of wings in tandem, that is, one set behind the other. When the day for the trial flight came, Oct. 14, 1903, public interest was aroused to a high pitch. The trial flight, however, was a failure, the machine falling into the Potomac river soon after the take-off, and with its fall there went Langley's hopes for success.

154. The Wright Brothers. And now we come to the thrilling story of the Wright brothers, Wilbur and Orville, two young bicycle mechanics of Dayton, Ohio. They began by experimenting with gliders, huge rectangular, kite shaped bodies, in the use of which they served their bird apprenticeship. In this particular field they owed much to Otto Lilienthal, a German, who before the Wright brothers began their work had made many daring flights in motorless planes, or gliders. Before attempting to build a power plane, the Wrights mastered the art of flying in their gliders, in the operation of which they learned many valuable lessons as to the principles of



Wilbur and Orville Wright

stabilization and control. The most important discovery which they made—a discovery, it should be noted, that made aviation finally possible—was the application of the principle of lateral balance. Langley and his European contemporaries had worked on the principle of rudder guidance for right and left motion, and elevator guidance for up and down motion. Stability as to the vertical sway of the plane, however, had not been solved. The Wrights solved this problem by the principle of *wing warping*. They discovered in their experiments with gliders that a plane could be stabilized by warping, that is, by bending the wings. They built their own plane. They designed and constructed their engine, a gasoline motor of 12 horse power. And on December 17, 1903, just two months after Langley's disastrous experiment with his Aerodome, they made their first successful flight near Kitty Hawk, N. C. On this occasion, for the first time in the history of man's long struggle to conquer the air, a heavier-than-air machine, carrying its own pilot, rose into the air and completed a short but successful flight. Four flights in all were made on this occasion, in one of which the plane rose of its own power and traveled a distance of 850 feet. On October 5, 1905, Wilbur Wright made a flight of 24 miles, during which the plane was under complete control, being brought back to the starting point at the close of the flight.

The story of the early accomplishments of the Wright brothers would not be complete without a brief mention of their sister Katharine. While Wilbur and Orville were spending every available hour in the prosecution of their experiment, Katharine, who was as much interested as they in their project, taught school and each month gave her money for the purchase of experimental supplies, while nightly she sewed the fabric to cover the wing frames.

From Icarus to Darius Green, and from Cayley to Langley, the record of man's attempt to master aerial flight had been one long story of tragic failure. But with the Wrights' success the spell was broken. Successful navigation of the air had to wait for the development of the modern light and powerful gas engine. It is significant to note that the gas engine and the airplane appeared almost simultaneously. In the light of present day motor production, the Wrights' engine was a very crude affair. It possessed sufficient power per pound of weight, however, to carry their plane for a short distance through the air and thus to demonstrate that with improved flying gear and engine equipment prolonged flight was possible.

155. Airplane Parts. The essential parts of an airplane may be classified into five major groups, namely, the body of the plane or fuselage, the power plant, the wings, and the stabilizing and controlling surfaces, Fig. 129.

Power plant. The power plant consists of the engine and the propeller, the blades of which are attached to the engine shaft. In this connection it will be of interest to make a comparison between the engines used by the Wright brothers and that which drove Lindbergh's plane across the ocean.

	Wright brothers'	*Lindbergh's*
Total horsepower	12 H.P.	225 H.P.
Weight of engine per H. P.	13 lb.	2.5 lb.
Total weight of engine, fuel, and oil	175 lb.	3400 lb.

Experience in aviation teaches us that an engine output of one horse power will support a gross weight of about 25 pounds. That is 1 H. P., 25 pounds; 100 H. P., 2500, and so on. Byrd's plane, in which he flew to the North Pole, had a gross weight—plane, engine, fuel, oil and pilots—of 22.6 pounds per H. P.; Lindbergh's, 23.6

pounds per H. P. Both planes when fully equipped for long flight were loaded dangerously close to the limit of 25 pounds per H. P.

Wings. An airplane is supported in its flight by the reaction of the air against the wings, which are set at a slight angle to the horizontal axis. Better lifting effect is obtained from a "cambered" wing than from a flat one. A cambered wing is one that is slightly curved from front to back.

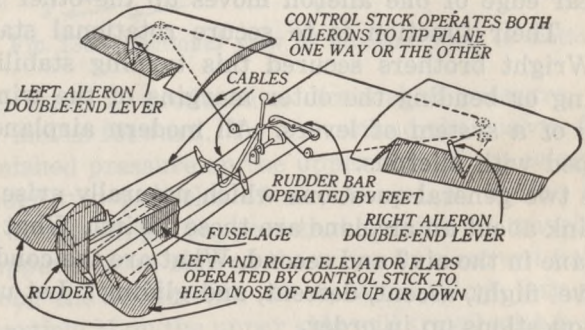


FIG. 129.—Airplane parts

Controlling and stabilizing devices. (a) Right and left motion through a vertical axis is secured by means of the rudder, which is controlled by small steel cables attached to foot levers in the cockpit, the place where the aviator sits. (b) Up and down motion through a horizontal axis is brought about by a corresponding up and down motion of the elevators, auxiliary horizontal tail wings. The elevators are attached to the rear of the horizontal stabilizing wings, which are placed near the tail end of the plane. The vertical controlling lever in the cockpit is called the "stick." The up and down motion of the elevators is accomplished by a forward and backward motion of the stick. (c) When the direction of motion changes from

straight line flight to a curved line, that is, when the aviator turns his plane, there is a tendency for the plane to roll or rotate about a longitudinal axis. To overcome this rolling tendency it is necessary for the aviator to "bank" his plane, that is to tilt it so that the outer wing system is higher than the inner. This is accomplished by means of the *ailerons*. These are auxiliary wings attached to the rear edge of the main wings. The ailerons are controlled by a sidewise motion of the stick. When the rear edge of one aileron moves up the other moves down. Their function is to secure rotational stability. The Wright brothers secured this turning stability by warping or bending the outer margins of the wings by means of a system of levers. All modern airplanes are equipped with ailerons.

The two general questions which naturally arise when we think about an airplane are these: First, What keeps the plane in the air? and second, What are the conditions for level flight, ascent, descent, and gliding? Let us take these questions up in order.

156. What Keeps an Airplane in the Air? An airplane is supported by the reaction of the air against the wings. In order to secure this reaction the plane must, of course, be in motion. To get at the answer to the question, What keeps an airplane in the air? let us consider the sectional view of the cambered wing shown in Fig. 130.

According to modern convention the direction of motion of the plane is represented by a feathered arrow. The angle which the lower face of the wing makes with a longitudinal axis is called the angle of incidence or "angle of attack." Because of the inclination of the wing the air is compressed on the under side, thus giving rise to an upward force. An upward thrust of this sort may be felt by swinging a card, which is tilted at an angle, through the air, or better still, by moving the open hand

rapidly through the water, the palm being held at an angle to the direction of motion. In the second place, on account of the curved surface of the upper side of the

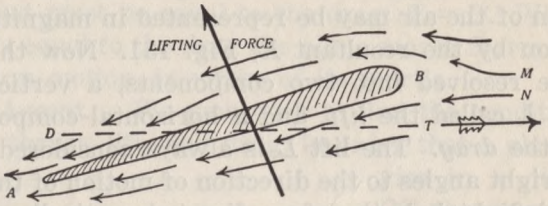


FIG. 130.—Sectional view of airplane wing in motion

wing, there is produced a diminished pressure, as the wing moves forward. To understand the reason for this diminished pressure on the upper side of the wing, let us consider two particles of air, *M* and *N*, at the leading edge of the wing. Suppose that particle *M* travels along the upper curved side of the wing in a curved path, and *N* along the under side. Now, other things being equal, the particle *M* on the upper side will travel with a greater speed than *N*, because it has farther to go in a given time. But we have learned (Art. 72) that when the speed of an air column is increased, the lateral pressure is diminished. The pressure on the upper side of the wing thus is much less than on the lower side. We then have, therefore, due to the motion of the plane, an increase of pressure on the under side of the wing and a diminution of pressure on the upper side. It is true that there are other subsidiary air reactions, but we shall not consider them here.

We may then say that an airplane is supported by the reaction of the air against the wings, there being an increase of pressure on the under side and a diminution of pressure on the upper side, resulting in a sustaining upward force, represented by the line *R* in Fig. 130. *It*

is estimated that about one fourth of the support comes from the compression below the wing and three fourths from the reduced pressure above.

157. Level Flight. As the wing moves forward, the reaction of the air may be represented in magnitude and direction by the resultant *R*, Fig. 131. Now this force may be resolved into two components, a vertical component *L* called the *lift*, and a horizontal component *D*, called the *drag*. The lift *L* is always considered as acting at right angles to the direction of motion of the plane. In level flight it is therefore directed vertically upward.

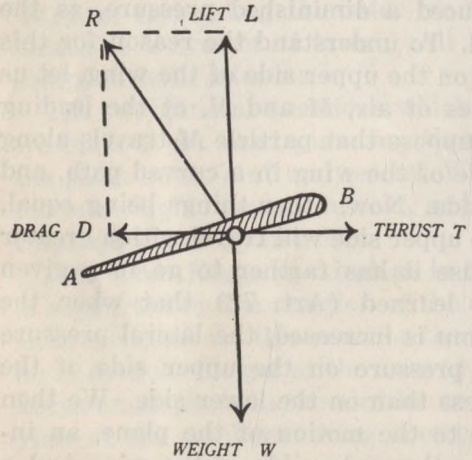


FIG. 131.—Conditions for level flight

The forward pull exerted upon the plane by the propeller is called the *thrust*, *T*. The weight of the plane is represented by the line *W*. It is assumed that all the forces in question act through the center of gravity of the system. It is clear that when the lift *L* is equal and opposite to the weight *W* that the plane will be sustained in the air. The fundamental equations of lift and drag may be written as follows:

$$lift = K_1 \times wing\ area \times velocity\ squared, \text{ that is}$$
$$L = K_1\ A\ V^2 \text{ and}$$
$$drag = K_d \times wing\ area \times velocity\ squared, \text{ or}$$
$$D = K_d\ A\ V^2$$

in which K_1 and K_d are the coefficients of lift and drag

respectively. The values of these coefficients are determined experimentally.

The conditions for level flight then are that the lift shall be equal to the weight, that is $L = W$, and also that the thrust must be equal to the drag, $T = D$. When the thrust is equal to the drag the plane moves forward with a uniform motion, in virtue of its inertia.

158. Ascent. We now wish to find the conditions of weight, thrust, lift, and drag, in order that a plane may rise in a direction OT, that is at an angle TOH, Fig. 132. We have here two resultant forces, W and R, both of which may be resolved into components. Let us take the weight W first. Completing the parallelogram we have the components W_1, acting downward at right angles to the direction of motion, and W_2, acting in the direction OW_2. Now at O erect a line OL representing the lift, acting at right angles to the direction of motion. The lift L is equal and opposite to the weight component W_1. Let D be the drag. Completing the parallelogram about L and D, we have the reaction R.

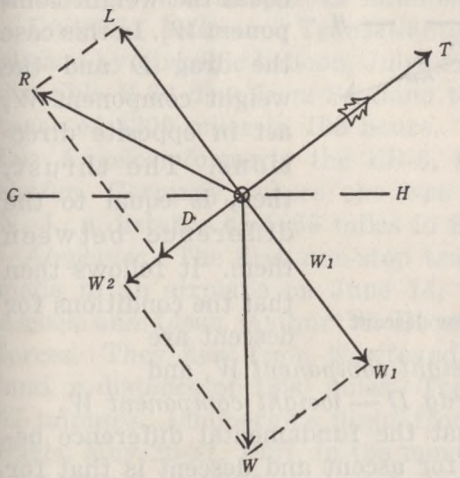


FIG. 132.—Conditions for ascent

Now it is evident that the weight component W_2 and the drag D are pulling back on the plane. The thrust T, therefore, will have to be a force equal to the sum of these two. That is to say, the conditions for ascent are

lift L = *component* W_1, and
thrust T = *drag* D + *weight component* W_2

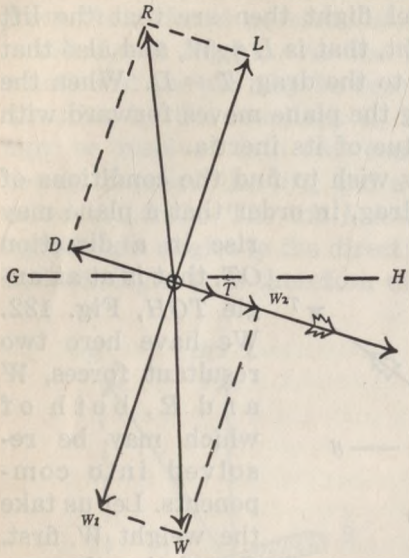


FIG. 133.—Conditions for descent

159. Descent. The plane is moving downward in the direction DT, Fig. 133. As before, W is the weight and R the reaction. Again, the lift L must equal the weight component W_1. In this case the drag D and the weight component W_2 act in opposite directions. The thrust, then, is equal to the difference between them. It follows then that the conditions for descent are

lift L = *weight component* W_1 and
thrust T = *drag* D — *weight component* W_2

It will be noted that the fundamental difference between the conditions for ascent and descent is that for ascent $T = D + W_2$, while for descent $T = D - W_2$.

When an airplane with power shut off descends without acceleration or rotation, it is said to *glide*. The angle of glide must be such that the lift L is equal to the weight component W_1, and the drag D must be equal to the weight component W_2. These conditions for gliding may be expressed as follows:

thrust $T = 0$,
lift L = *weight component* W_1, and
drag D = *weight component* W_2

160. Famous Atlantic Crossings. *Sailing vessels.* On August 3, 1492, Christopher Columbus set sail from Palos, Spain, on his voyage of discovery. He reached San Salvador Island 70 days later. On June 15, 1859, the Dreadnaught, an American clipper ship, made the passage from New York to Cape Clear, Ireland, in 12 days For a sailing vessel this was a record trip.

Steamships. The big ocean liner S.S. Bremen, July, 1929, steamed from Cherbourg, France, to New York in 4 days, 17 hours, and 42 minutes.

Dirigible balloons. Transatlantic flight was accomplished by dirigible balloon, July, 1919, when the British dirigible R-34 flew from Scotland to Long Island, a distance of 3200 miles in 108 hours. In 1924 the dirigible Los Angeles, formerly the ZR-3, flew from Friedrichshafen, Germany, where she was built, to Lakehurst, N. J., a distance of 5066 miles in 81 hours, 17 minutes.

Airplanes. The first non-stop transatlantic flight was made in an airplane on June 14, 1919, by Capt. John Alcock and Lieut. Arthur W. Brown, of the British air forces. They flew from Newfoundland to Clifden, Ireland, a distance of 1890 miles. The time was 15 hours, 57 minutes. Lindbergh's flight from New York City to Paris, May 20-21, 1927, in the monoplane, "Spirit of St. Louis," was one of the most daring and spectacular events of modern times. His route from New York City to Paris was almost a direct line, as measured on the surface of the earth as a sphere. On a flat map his course appears to be distorted. If by means of a string the distance is measured on the surface of a globe, it will be seen that the course he took, by way of Nova Scotia, Newfoundland, and the southern coast of Ireland, is the shortest distance between his starting and stopping points. He travelled the distance, 3610 miles, in 33 hours and 29 minutes.

CURVILINEAR MOTION

161. Centrifugal Force. We shall now devote our attention to curvilinear motion, and some of its practical applications. When a body moves in a curved path, it tends at every instant to fly off in a straight line, in obedience to the first law of motion. This tendency of a body executing curvilinear motion to travel in a straight line causes it to exert a force which is directed outward from the center of rotation. This outwardly directed force is centrifugal force. The presence of such a force in the case of a body rotating about a central point may be demonstrated by attaching a metal ball to a string and whirling the ball around in a vertical plane. A distinct pull on the string may be felt, becoming greater as the speed of rotation increases. The outward force exerted by the rotating ball is called *centrifugal force;* the reaction exerted upon the ball by the string toward the center of rotation is called *centripetal force.*

Centrifugal force is exerted *by* the rotating body and is directed from the center toward the circumference; centripetal force is exerted *on* the body and is directed from the circumference toward the center. These forces are equivalent to action and reaction; they are equal to each other in magnitude, and are oppositely directed.

Many illustrations of the effect of centrifugal force may be seen in the ordinary affairs of everyday life. For example, the mud and water on bicycle or automobile wheels are thrown off, due to the centrifugal force exerted by the rotating bodies when these vehicles are in rapid motion. A freshly oiled typewriter sometimes spatters oil from the type bars when the machine is operated. A bicycle rider in going around a corner instinctively leans inward to overcome the centrifugal force which tends to overturn him. For the same reason the circus

rider leans in toward the center of the ring. Likewise, on railway curves the outer rail is laid higher than the inner rail to overcome the centrifugal force exerted by the train. Also, highways are "banked" around curves, that is, the outer side of the roadbed is made higher than the inner side, to counteract the overturning effect of centrifugal force. Other examples of centrifugal and centripetal forces will be discussed in detail in the topics which follow.

162. The Laws of Centrifugal Force. Before proceeding with a further study of applications of the principle of centrifugal force, however, we shall consider the laws which govern centrifugal and centripetal forces. These laws may be stated as follows:

I. *Centrifugal (also centripetal) force is directly proportional to the mass of the rotating body.* That is, the greater the mass the greater the force. For example, if a mass of 1 pound is whirled around with a given speed, a given centrifugal force (pull on the string) will be exerted; if, now, a mass of 10 pounds be rotated with the same speed, the centrifugal force will be 10 times as great as in the first case.

II. *Centrifugal force is directly proportional to the square of the speed.* If the speed of the rotating body be increased from 1 foot per second to 10 feet per second, the centrifugal force will be increased from 1 to 10^2, that is from 1 to 100.

III. *Centrifugal force, for a given speed, is inversely proportional to the radius of rotation.*

The three laws may be expressed by means of the equation,

$$centrifugal\ force = \frac{mass \times speed\ squared}{radius}$$

$$F = \frac{mv^2}{r}$$

This equation expresses the centrifugal force F in *absolute units,* that is, poundals or dynes, as the case may be. To change poundals to pounds of force, divide by 32; dynes to grams of force, divide by 980.

Example. A mass of 6 pounds is whirled around at the end of a string 2 feet in length with a speed of 4 feet per second. Find the pull on the string due to rotation in (a) poundals; (b) pounds. *Solution:*

$$F = \frac{mv^2}{r} = \frac{6 \times 4^2}{2} = 48 \text{ poundals}$$

$$= \frac{48}{32} = 1.5 \text{ pounds}$$

EXERCISE. **43.** A mass of 98 grams attached to the end of a string 20 cm in length is whirled around with a speed of 20 cm per second. Find the centrifugal force exerted by the rotating body in (a) dynes; (b) grams of force.

163. Experiments Illustrating Centrifugal Force. Experiments with whirling table. (a) If some mercury and water be placed in a receptacle and rotated rapidly, the mercury will take a position in the equatorial region of the globe, with a zone of water on either side, Fig. 134. The mercury, having a greater mass per unit of volume

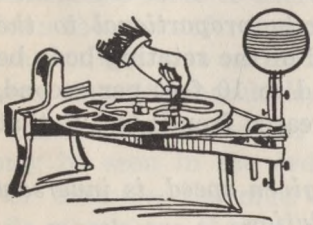


FIG. 134



FIG. 135

Illustrating centrifugal force

than that of the water, exerts a greater centrifugal force, and hence gets farthest from the axis of rotation. This experiment illustrates the principle of the modern cream separator, by means of which the cream is separated from the milk.

(b) If a circular band of metal be whirled about a central vertical axis, Fig. 135, it will flatten at the poles, due to the fact that the particles in the region of the equator have a greater speed, and hence exert a greater centrifugal force than do the particles in the region of the poles.

(c) A striking example of the effect of centrifugal force may be given by whirling a small bucket of water around a vertical plane. In this case the water is kept in the bucket by the centrifugal force tending to throw it away from the center of rotation.

Many practical applications of centrifugal force are in commercial use at the present time. Centrifugal machines are employed in manufacturing plants for the mechanical drying of salt, sugar, and chemical products of a similar nature, and in laundries for the drying of clothes. Other applications of the centrifugal principle are discussed in the topics which follow.

164. The Cream Separator. Two important practical applications of the principle of centrifugal force are those which have to do with the modern cream separator, and the Babcock milk tester.

Fresh milk consists of fats (cream) and skim milk (water, casein, and sugar) intimately mixed. The cream occurs in the milk in finely divided particles. If fresh milk is allowed to stand for several hours in shallow pans, the cream particles, being slightly lighter than the milk particles, slowly rise to the surface where the cream may be skimmed off. The gravity, or shallow pan method, which was in common use up to the time of the invention of the centrifugal separator, was a very slow and relatively inefficient process. In gravity separation and skimming, as usually practiced, 5 per cent or more of the cream remained in the milk; with a good centrifugal separator 98 to 99 per cent of the butter fat is obtained. This latter process also has an advantage in that it is possible to separate the cream quickly and as soon as the milk is taken from the cows. One type of the cream separator is shown in Fig. 136.

The essential parts of a cream separator are a bowl, which is

capable of rotating at high speed, surrounded by a stationary jacket, and a gearing device for rotating the shaft connected to the bowl. The modern separator contains a set of discs within the bowl. When in operation the discs and bowl make from 6000 to 8000 revolutions per minute. In Fig. 137 there is shown

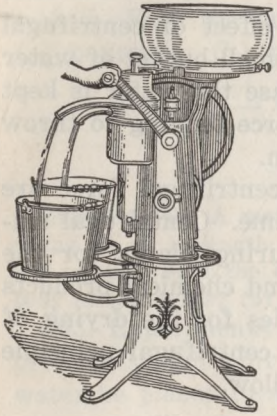


FIG. 136

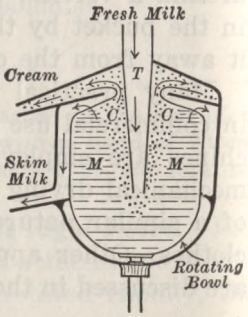


FIG. 137

Cream separator

a section of the separating chamber, the discs being omitted in order that the essential principle may be made clear. The bowl is rapidly rotated by means of a "worm" gear which is not shown. Fresh milk enters through the tube *T*, and soon acquires the rotary motion of the bowl. Both the cream and the milk are thrown outward from the central axis of rotation by centrifugal force. But the milk particles *M* being heavier than the cream particles of the same size experience the greater force and soon occupy the outer section of the bowl, the cream particles, *CC*, being nearer the center. The skim milk escapes from the bowl through holes at the top near the circumference, while the lighter cream escapes through holes nearer the axis of rotation.

165. The Babcock Milk Tester. This mechanical device is nothing more than a simple form of cream separator in which the milk is placed in a series of two or more specially prepared and graduated test tubes, Fig. 138. The milk to be tested is placed in

the tubes, which are then whirled rapidly about a central axis. The tube supports are pivoted so as to allow the tubes to swing outward on rotation. When the axis is rotated, the tubes assume a horizontal position, the skim milk appearing in the outer part (bottom of the tube), and the cream at the top. By means of the scale which is etched upon each tube, it is possible at once to determine the amount of butter fat (cream) present in the milk.

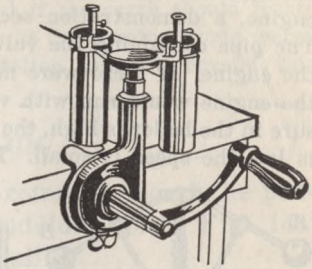


FIG. 138.—Two-tube Babcock milk tester

This testing device was invented by Professor S. M. Babcock and was offered by him without royalty charges to the dairy farmers of the country. By means of the Babcock test it is possible for any dairyman easily and accurately to determine the butter fat producing qualities of each member of his herd.

166. The Centrifugal Pump. The *centrifugal pump* is extensively used in pumping water in city systems. A front view and

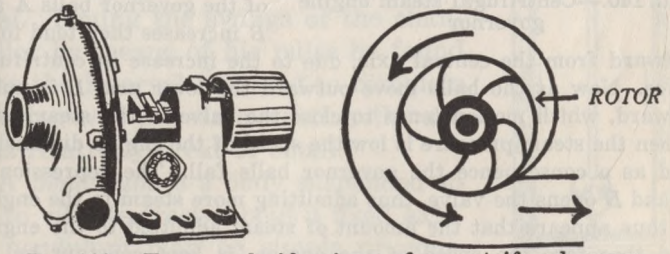


FIG. 139.—Front and side views of a centrifugal pump

also a sectional view of such a pump is shown in Fig. 139. The driving part of the pump consists of a wheel, called the rotor, having curved fan-like blades. The water enters the pump at the axis of the rotor and is discharged through the delivery pipe with great force by the rotating blades.

Centrifugal pumps are also used to circulate the water in automobile cooling systems. Machines of the centrifugal type called blowers are largely employed to ventilate buildings, and to furnish a forced draft to furnaces.

167. The Centrifugal Governor. A practical application of centrifugal force is that exemplified by the "governor" of the steam engine, a demonstration section of which is shown in Fig. 140. The pipe containing the valve *V* admits steam from the boiler to the engine. If there were no regulating valve in this intake pipe the engine would run with variable speed; that is, when the pressure in the boiler is high, the speed is great, and when the pressure is low the speed is small. The function of the governor is to so regulate the amount of steam admitted to the engine as to give uniform motion to the fly wheel. The flow of steam is controlled by the turning of the valve *V*. The balls *A* and *B* are caused to rotate about a vertical axis by means of a drive belt from the engine. When the steam pressure is high the engine speeds up. As the rotation of the governor balls *A* and *B* increases they tend to fly outward from the central axis, due to the increase in centrifugal force. Now as the balls move outward the lever arm *L* is pulled upward, which motion tends to close the valve in the steam pipe. When the steam pressure is low the speed of the engine diminishes, and as a consequence the governor balls fall. The depression of *A* and *B* opens the valve, thus admitting more steam to the engine. It thus appears that the amount of steam admitted to the engine, and therefore the speed of the engine, is kept constant by the operation of the governor balls.

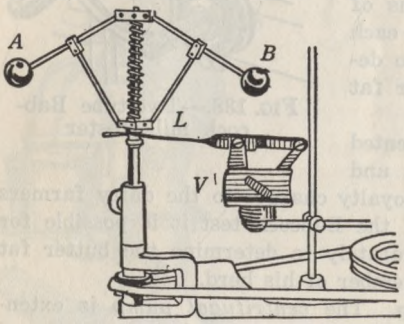


FIG. 140.—Centrifugal steam engine governor

The centrifugal governor principle is also used in the automobile speedometer.

168. Centrifugal Force and Weight. There are two reasons why a given body weighs less at the equator than at the poles. The first is because the force of gravity, as has already been explained, is less at the equator than at the poles; and the second is that the centrifugal force at the equator is greater than at or near the poles. Bodies at the equator have a speed, due to the rotation of the earth, of more than 1000 miles per hour. The resulting cen-

trifugal force is about 1/289 of the force of gravity. Now, since centrifugal force is proportional to the square of the speed, and the square of 17 is 289, it follows that if the earth should rotate 17 times as fast as it now does, bodies at the equator would weigh nothing; and should the rate of rotation increase beyond this value, they would fly off into space.

STOP for TEST

The Pendulum

169. Introductory. The most common use of the pendulum is exemplified by the pendulum clock, Fig. 141, in the measurement of time. Galileo was one of the first to investigate and establish the laws of the pendulum. It is related that he discovered the fact that the vibrations of a pendulum are performed in equal intervals of time, that is to say the vibrations are *isochronous,* by observing the vibrations of a swinging chandelier in a cathedral. Timing the swings of the chandelier by means of his pulse he found that the interval required to complete a swing was the same whether the arc was relatively great or small.



Fig. 141
Laboratory clock

A *pendulum* is a body suspended so as to vibrate freely about a fixed axis. A pendulum may be simple or compound. A *simple pendulum* (sometimes called a mathematical pendulum) is considered to be a material particle suspended by a weightless thread. Of course no such pendulum actually exists, since any thread, however small, must have some weight. A small lead bullet, however, suspended by a fine thread may be considered as an illustration of a simple pendulum. A *compound pendulum* is any body suspended so as to vibrate freely. A meter

stick vibrating about one end is a compound pendulum. In fact all pendulums in actual use are compound pendulums, differing only in their approximation to the ideal simple type. The pendulum of a clock, for instance, is a compound pendulum in which most of the mass is in the bob.

170. Period, Amplitude, Frequency. A study of the pendulum is important because it serves to give the student a clear idea of the meaning of *period, amplitude,* and *frequency,* terms which are used later on in our study of wave motion as relating to sound, light, and electricity. Let us begin by considering the vibratory motion of a pendulum, such as is shown in Fig. 142, as it swings back and forth through the arc *ABC*.

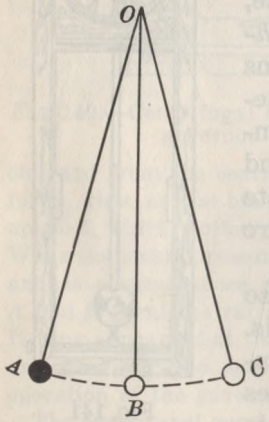


FIG. 142

A *single vibration,* sometimes called a *simple vibration,* is the swing from one side of the arc to the other, that is from *A* to *C*. A *complete* or *double vibration* is a swing from *A* to *C* and back to *A* again. The *period* of vibration is the time in seconds required to make one vibration. The period of a single vibration is usually designated by the letter t; that of a complete vibration by the letter T. When we speak of the period of sound, light, or electrical waves it is customary to consider complete vibrations. At this stage of our study, however, it will be simpler to confine our attention to simple vibrations. It is important to note that the term *period* refers to the time in seconds required to make one vibration and has no reference to the time required for the swinging pendulum to come to rest.

Amplitude is one-half the arc described by a single swing, as for example from A to B or from B to C. *Frequency* is the number of vibrations per second, and is usually designated by the letter n.

The period of a vibrating body is equal to the reciprocal of its frequency. That this is the case may be seen from a consideration of the following example. If a pendulum makes 20 simple vibrations in 10 seconds, its frequency (number of vibrations per second) is $n = 20/10 = 2$. But its period (time of one vibration) is $t = 10/20 = \frac{1}{2}$. Hence it appears that

$$period = \frac{1}{frequency}$$

$$t = \frac{1}{n}$$

171. Motion of the Pendulum Explained. Consider the pendulum, Fig. 143, to be drawn aside to the point A. It is acted upon by the force of gravity represented by the vertical line AG. This force may be resolved into two components, one, AC, producing a tension in the string, and the other, AB, tending to produce motion toward the position of rest at D. The pendulum then tends to move from A to D due to a component of the force of gravity. It continues to swing beyond D because of its inertia. If there were no friction to interfere, the pendulum bob would rise as high on one side of the arc as on the other, that is to say, it would never come to

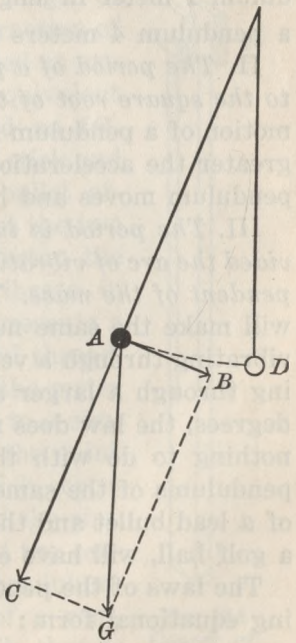


FIG. 143

rest. The friction which brings the swinging pendulum to rest is that which it encounters at the point of suspension and that due to the resistance of the air.

172. Laws of the Pendulum. When we consider the vibration of a pendulum, the question which naturally arises is this: What is the relation between the period of vibration and (a) its length, (b) the acceleration of gravity g, (c) its amplitude, and (d) its mass? It may be shown both experimentally and mathematically that the following relations hold, as expressed by the laws of the pendulum:

I. *The period of a pendulum is directly proportional to the square root of its length.* If the period of a pendulum 1 meter in length is 1 second, then the period of a pendulum 4 meters in length will be $\sqrt{4} = 2$ seconds.

II. *The period of a pendulum is inversely proportional to the square root of the acceleration of gravity, g.* The motion of a pendulum is due to the force of gravity. The greater the acceleration of gravity the more quickly the pendulum moves and hence the shorter its period.

III. *The period is independent of the amplitude (provided the arc of vibration is relatively short), and is independent of the mass.* This means that a given pendulum will make the same number of swings per second when vibrating through a very small arc as it will when swinging through a larger arc. If the arc is greater than 10 degrees, the law does not hold. The mass of the bob has nothing to do with the period of the pendulum. Two pendulums of the same length, the bob of one consisting of a lead bullet and the other a mass of lead as large as a golf ball, will have equal periods of vibration.

The laws of the pendulum may be stated in the following equational form:

$$t = \pi \sqrt{\frac{l}{g}}$$

in which t is the period of a *simple* vibration in seconds; l is the length of the equivalent simple pendulum in feet or centimeters; g is the acceleration of gravity in feet or in centimeters per second per second. The constant π (3.1416) occurs here because the mathematical derivation of this equation involves the ratio of the circumference of a circle to its diameter (circumference $= 2\pi$).

The equation for the period of a *complete* vibration is

$$T = 2\pi \sqrt{\frac{l}{g}}$$

173. Center of Oscillation. Having given the length of a compound pendulum, it is important to know how to find the length of an equivalent simple pendulum, since l in the equation of Art. 172 is the length of a simple pendulum. To find what fraction of the length of a meter stick suspended at one end is equal to the length of an equivalent simple pendulum, we may proceed as follows: Suspend side by side a meter stick and a pendulum consisting of a lead bullet attached to a fine thread, Fig. 144. Set the two pendulums vibrating, and then shorten the simple pendulum until the two vibrate in the same time. The meter stick represents a compound pendulum; the bullet and thread, a simple pendulum. That length of the meter stick, which is equivalent in length to a simple pendulum, is the distance from the point of suspension S to the point C, which corresponds to the position of the bob of the simple pendulum. The point C is called the *center of oscillation* of the compound pendulum. When we speak of the length l of a compound pendulum we mean a length that is equivalent to a simple pendulum having the

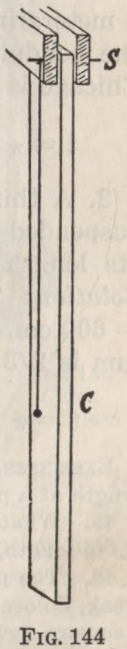


FIG. 144
Center of oscillation

same period. If the compound pendulum be a thin uniform rod, as in the case of the meter stick; and, if suspended from the end, l is 2/3 of the entire length of the meter stick. If the pendulum is not uniform, however, as in the case of a ball bat, the length l has to be determined by experiment.

The center of oscillation of a compound pendulum is sometimes called the *center of percussion.* This is the point at which the most effective blow can be struck. The center of percussion of a hammer is in the head; the center of percussion of a ball bat is at a point near the large end.

Examples. 1. Find the period of a simple pendulum 4 meters in length, at Chicago. *Solution*: The length of the pendulum is 4 m = 400 cm, and the value of g at Chicago is 980.26 (Table X, Appendix). Then

$$t = \pi \sqrt{\frac{l}{g}} = 3.1416 \sqrt{\frac{400}{980.26}} = 2.007 \text{ sec.}$$

2. A thin steel rod of uniform cross sectional area is suspended at one end and set in vibration at Chicago. Its length is 6 meters. Find its period of vibration. *Solution*: The length of this compound pendulum is 6 m = 600 cm, and the length of an equivalent simple pendulum is 2/3 of 600 = 400 cm. Then

$$t = \pi \sqrt{\frac{l}{g}} = 3.1416 \sqrt{\frac{400}{980.26}} = 2.007 \text{ sec.}$$

EXERCISES. **44.** Find the period of a simple pendulum 6.4 ft. in length at a place where g = 32 ft. per second per second.

45. What is the period of vibration of a thin uniform rod 12 ft. in length, the value of g being 32 ft. per second per second?

46. The period of a simple pendulum 1 m in length on Pike's Peak, Colorado, is 1.004 seconds. Find the value of g at that place. Compare your result with that given for g, Pike's Peak, Table IX, Appendix.

174. Mechanics of the Clock Pendulum. Since the vi-

brations of a pendulum are performed in equal intervals of time, that is, they are isochronous, all that is needed in order to use the pendulum as a recorder of time, is some mechanical device to keep the pendulum vibrating and to enable it to regulate the motion of the hands. The motive power of the clock is supplied by weights or springs. The motion of the hands is regulated by means of an escapement *C*, Fig. 145. As the pendulum swings to and fro the projections of the escapement catch alternately in the teeth of the escapement wheel, *D*, thus allowing only one tooth to escape for each double swing. The slight pushes communicated to *C* by the teeth of *D* as they alternately catch and escape keep the pendulum in motion. If the escapement wheel has 30 teeth, it will rotate once while the pendulum makes 30 complete vibrations.

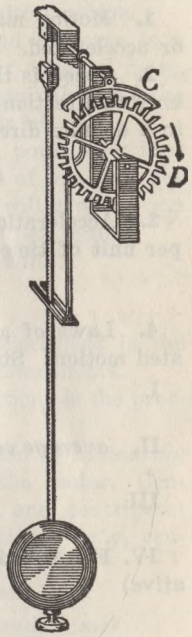


FIG. 145
Clock pendulum

In order that the periods shall be equal, the length of the pendulum must always be the same. In summer, however, the pendulum lengthens, and hence the clock tends to lose time. In winter the pendulum shortens and the clock gains time. Corrections must therefore be made for changes in length due to changes in temperature. This may be done by moving the bob up or down by means of a nut or thumb screw. Changes in the length of a pendulum due to changes in temperature may also be automatically corrected by means of compensating devices. A *compensating pendulum* is one that is made of two or more metals so regulated as to expand in opposite

directions, and thereby keep the length of the pendulum constant.

SUMMARY

1. Motion may be (a) rectilinear or curvilinear; (b) uniform or accelerated.

2. Speed is the rate of motion per unit of time, without reference to direction; velocity is the time rate of motion with reference to a definite direction. In both cases

$$speed,\ velocity = \frac{space}{time} = \frac{s}{t}$$

3. Acceleration is the increase or decrease (change) of velocity per unit of time; that is,

$$acceleration = \frac{change\ of\ velocity}{time}$$

4. Laws of acceleration for a body having uniformly accelerated motion: Starting from rest

I. $$velocity = acceleration \times time$$
$$v = at$$

II. $$average\ velocity = \tfrac{1}{2}\ (vel.\ at\ beginning + vel.\ at\ end)$$
$$= \tfrac{1}{2}\ (v' + v)$$

III. $$space = average\ velocity \times time$$
$$s = \tfrac{1}{2}\ at^2$$

IV. For a body having an initial velocity v' (positive or negative)

$$v = v' \pm at, \text{ and}$$
$$s = v't \pm \tfrac{1}{2}\ at^2$$

5. The acceleration of gravity g is usually taken as 32 ft. per second per second, or 980 cm per second per second. This is sometimes written $g = 32\ ft./sec.^2 = 980 cm/sec.^2$

6. Inertia is the resistance which a body offers to a change in its condition of rest or motion.

7. Momentum = mass × velocity = mv.

8. Impulse = force × time = Ft.

9. Newton's laws of motion: I. Every body tends to preserve its state of rest or uniform motion in a straight line unless compelled by some external force to change that state.

II. Change of motion is proportional to the impressed force, and takes place in the direction in which the force is applied. "Change of motion" may be interpreted to mean *change of momentum* and

"impressed force" to mean *impulse*. The law then may be written

$$\textit{impulse} = \textit{change of momentum}$$

$$Ft = \textit{mass} \times \textit{change of velocity}, \text{ hence}$$

$$F = m\,a$$

III. To every action there is an equal and opposite reaction.

10. Gravitational units of force: (a) A pound of force is a force equivalent to the attraction of the earth for a pound mass, at sea level. (b) A gram of force is a force equivalent to the attraction of the earth for a gram mass, at sea level.

11. Absolute units of force, $F' = ma$: (a) A poundal is that force that will give a pound mass an acceleration of one foot per second per second. (b) A dyne is that force that will give a gram mass an acceleration of one centimeter per second per second.

12. Relation of gravitational units to absolute units:

1 pound of force = 32 poundals

1 gram of force = 980 dynes

13. Composition of forces, velocities, and accelerations is the process of finding the resultant of two or more components.

14. Resolution of forces, velocities, and accelerations is the process of separating a resultant into its components.

15. Centrifugal force is the outward pull exerted by a rotating body. Centripetal force is the reaction toward the center. Centrifugal force is exerted *by* the rotating body, and centripetal force is exerted *upon* the body. For a given rotating body, centrifugal force is numerically equal to centripetal force.

16. Laws of centrifugal and centripetal forces:

$$\textit{centrifugal force} = \frac{\textit{mass} \times \textit{velocity squared}}{\textit{radius}} = \frac{mv^2}{r}$$

This equation gives results in absolute units; that is, poundals or dynes, depending on the units in which m, v, and r are expressed.

17. Simple and compound pendulums: (a) A simple pendulum is a material body suspended by a thread of negligible weight so as to vibrate freely. (b) A compound pendulum is a body having appreciable mass throughout suspended so as to vibrate about an axis of support.

18. Period of vibration: The period t of a simple or single vibration is the time in seconds required to swing from one side of the arc to the other. The period T of a complete or double vibration is the time required to swing from one side of the arc to the other and back again. The period of a complete vibration is therefore twice that of a simple vibration.

19. Amplitude of vibration is one-half the arc swept out in a single vibration.

20. Frequency n is the number of vibrations per second.

21. Period $= \frac{1}{\text{frequency}}$, or frequency $= \frac{1}{\text{period}}$.

22. Laws of the simple pendulum: I. The period is directly proportional to the square root of the length.

II. The period is inversely proportional to the square root of the acceleration of gravity.

III. The period is independent of the amplitude of vibration (when the amplitude is not greater than 10 degrees) and the mass of the bob.

23. The laws of the pendulum may be expressed in equational form as

$$t = \pi \sqrt{\frac{l}{g}}, \text{ and } T = 2\pi \sqrt{\frac{l}{g}}$$

in which t is the period of a simple vibration and T the period of a complete vibration, l the length in centimeters or feet, and g the acceleration in cm/s/s or ft./s/s.

REVIEW EXERCISES

A

1. Explain what is meant by the statement that motion and rest are relative terms. Give illustrations.

2. Classify motion as to (a) direction, (b) rate.

3. Explain the difference between the meaning of the terms *speed* and *velocity*.

4. Define acceleration, and give two or more examples of accelerated motion.

5. What is the distinction between positive and negative acceleration?

6. Explain the meaning of each term in the equation $v = at$, and $s = \frac{1}{2}\, at^2$.

7. A body acted upon by a constant force starts from rest and at the end of the 1st second has a velocity of 5 ft. per second; at the end of the 2d second, 10 ft. per second, and so on. (a) What is its velocity at the end of 10 seconds? (b) Over what space will the body pass in 10 seconds?

8. Over what space does the body (Ex. 7) pass during the 10th second?

9. An automobile starts from rest with an acceleration of 4 ft. per sec. per sec. In what time will it acquire a speed of 30 mi. per hour?

10. What distance in feet will the automobile (Ex. 9) travel in the time required to attain a speed of 30 mi. per hour?

11. Define inertia and give two or more examples of inertia.

12. What is momentum? What are the two factors that determine the momentum which a body possesses?

13. State the first law of motion. Why is this law called the law of inertia?

14. State the second law of motion. What interpretation did Maxwell give to the terms "change of motion" and "impressed force?" What is impulse?

15. State the third law of motion. What is meant by the terms action and reaction?

16. Compare the momentum of a body having a mass of 2 ounces and a velocity of 40 ft. per second with that of a body having a mass of 2 lb. and a velocity of 5 ft. per second.

17. Explain each term of the equation $mv = m'v'$, and explain, also, to what law of motion it applies.

18. Explain the terms (a) composition of forces; (b) resolution of forces. To what other physical quantities may the principles of composition and resolution be applied?

19. Show by means of a sketch how you would find the resultant of two forces acting on a body at A, one of the forces AB being equal to 60 lb. and the other AC, 40 lb., the angle between the components AB and AC being 30 degrees.

20. Given the resultant of two forces and one of the components, to find the other force. The magnitude of the resultant is 120 units, and the given component 80 units. The angle between the resultant and the given component is 20 degrees. Show by means of a sketch how you would find the other component.

21. A rod 4 ft. in length is supported at the ends A and C. Where should a 50 lb. weight be placed so that the end A shall support 20 lb. of the weight and the end C 30 lb.?

22. Explain the distinction between *centrifugal force* and *centripetal force.*

23. State the laws of centrifugal force.

24. Why does a bicycle rider lean inward in going around a curve?

25. A street car is running around a curve. Which rail receives the thrust of the flanged wheels?

26. Why is the outer rail of a track laid higher than the inner rail?

27. Why does an aviator bank his plane in making a turn?

28. What is the difference between a simple pendulum and a compound pendulum?

29. Define amplitude of vibration, period, frequency.

30. What is the relation of period to frequency?

31. A pendulum makes 160 single vibrations in 4 minutes. What is (a) its period? (b) its frequency?

32. State the laws of the pendulum.

33. How is the period of a pendulum affected (a) when its length is increased? (b) when the force of gravity is increased?

B

34. A stone is thrown vertically downward from the top of a cliff with an initial velocity of 15 ft. per second. It strikes the ground at the bottom of the cliff in 5 seconds. How high is the cliff?

35. An automobile running at the rate of 30 mi. per hour is brought to rest within a space of 88 ft., the negative acceleration being uniform. What time in seconds was required to bring it to rest?

36. A ball is projected vertically upward with an initial velocity sufficient to cause it to rise to a height of 313.6 m. (a) What time will be required for the ball to fall through this distance? (b) How many seconds was it in the air from the time it started to rise until it returned?

37. What was the initial velocity of the ball (Ex. 36)?

38. A ball is fired from a cannon having a mass of 1 ton. The velocity of the ball is 2000 ft. per sec. The velocity of the gun's recoil is 2 ft. per sec. What is the mass of the ball?

39. Explain how a definition of force ($F = ma$) may be derived from the second law of motion.

40. What is the distinction between a *pound of force* and a *poundal*?

41. What is the distinction between a *gram of force* and a *dyne*?

42. A force of 128 lb. is equivalent to how many poundals? 25,480 dynes are equivalent to how many grams of force?

43. A 12 ft. plank spans a stream. A man weighing 168 lb. crosses on the plank. Call one end of the plank A and the other B, and assume that the man walks from A toward B. What part of his weight rests upon each support when he is (a) 4 ft. from A? (b) 6 ft. from A? (c) 8 ft. from A?

44. A boy draws a sled by means of a rope which makes an angle of 30° with the horizontal. He exerts on the rope a force of 30 lb. (a) What is the horizontal component of this force acting upon the sled? (b) the vertical component? (c) Make a drawing to illustrate the solution of this problem.

45. Explain the use of the following equation and state the meaning of each term contained therein: $F = mv^2/r$. Does this equation give results in absolute or gravitational units?

46. A mass of 16 lb. is whirling around in a circle at the end of a thin rod 4 ft. long with a speed of 20 ft. per second. What is the centrifugal force exerted by the mass, in pounds?

47. A mass of 4 lb. attached to the end of a thin rod AB is making 120 revolutions per minute (r. p. m.) about the end B. The centrifugal force is 30 lb. What is the length of the rod?

48. Explain the meaning of each term in the equations

$$\text{(a)}\quad t = \pi\sqrt{\frac{l}{g}}, \qquad \text{(b)}\quad T = 2\pi\sqrt{\frac{l}{g}}$$

C

49. A ball having a mass of 6 lb. starts from rest and rolls down an incline with an acceleration of 8 ft./s/s. What is its momentum at the end of the 4th second?

50. A stream which is a mile wide flows with an average velocity of 1 mi. per hour. A man is in a boat at A on one side of the stream. On the other side of the stream are three points, B, C, and D. B is directly across from A; C is one mile downstream from B; and D is one mile upstream from B. The man starts from A and rows at the rate of 1 mi. per hour. At the end of an hour where will he be if he directs his boat (a) directly upstream? (b) directly downstream? (c) directly across the stream?

51. By means of a sketch show where he will land on the opposite side of the stream (Ex. 50) if he directs his boat at the start toward D, which is 1 mi. upstream from B, and rows at the

rate of 1 mi. per hour until he reaches the opposite shore.

52. Under the conditions named in Ex. 50 can he possibly land at B? Explain.

53. Give the reasons why a body of given mass weighs less at the equator than at or near the poles of the earth. NOTE: Consult Arts. 10 and 168.

54. Explain how to find by experiment the center of oscillation of a compound pendulum.

55. It is related that at one time Thomas Jefferson proposed that the United States Government adopt as the legal standard of length a metal rod of uniform thickness and density which would beat seconds (period of a single vibration = 1 second) when suspended from one end. What would be the length of such a rod if the determinations were made at Washington, D. C.? Bear in mind that you are dealing here with a *compound* pendulum.

56. Explain how it is possible to determine the acceleration of gravity g for any given place by means of a pendulum of known period.

57. Explain the use of the following parts of an airplane: rudder, elevator, aileron.

58. Explain the meaning of the following terms: thrust, lift, drag.

59. What keeps an airplane in the air?

60. What are the conditions for level flight?

NOTE: For additional problems, see Appendix.

CHAPTER VI

MECHANICS OF SOLIDS (Continued)

Work, Power, Energy

175. Work. *Work* is the overcoming of resistance through space. As used in physics, the term work involves two ideas, *force* and *space*. If a man were to hold up a heavy weight without giving it motion, he would not be doing work in the physical sense. The moment he moves the body, that is, applies force through space, he does work. The equation for work is

$$work = force \times space$$
$$w = Fs$$

In the equation $w = Fs$, the space s is measured in the direction of the application of the force. For example, if a bucket of mortar having a weight (force) of 75 pounds be lifted vertically by means of a rope from A to B, Fig. 146, a distance of 12 feet, the work performed would be $75 \times 12 = 900$ foot pounds. Suppose now that the load (75 pounds) is carried up the ladder to B, a distance of 13.5 feet. While the space traversed by the workman in going up the ladder is 13.5 feet, yet the space s through which he actually exerts the force of 75 pounds against gravity is, as before, 12 feet, and the work done is again 900 foot pounds.



FIG. 146

Work does not involve the factor of time. If a weight be lifted from the floor

to the table in 10 seconds, a given amount of work is done. Again, if the same weight be lifted through the same space in 10 minutes, or in 10 hours, the same amount of work is done as in the first case. A train of cars hauled from one station to another requires a given amount of work, no matter what time is involved.

176. Units of Work. Since work is force times space, the units in which work is measured depend upon the units chosen for force and space. There are, therefore, for work, as in the case of force, two sets of units, the gravitational and absolute.

Units of work { *gravitational* { *foot pound*, *gram centimeter*; *absolute* { *foot poundal*, *erg*

The *foot pound* is the work done by a force of one pound acting through a space of one foot. The *gram centimeter* is the work done by a force of one gram acting through a space of one centimeter. The kilogram meter is sometimes used instead of the smaller unit, the gram centimeter. A *kilogram meter* is the work done by a force of one kilogram acting through a space of one meter.

The *foot poundal* is the work done by a force of one poundal acting through a space of one foot. The *erg*, derived from the Greek *ergon* for work, is the work done by a force of one dyne acting through a space of one centimeter.

The two units of work ordinarily employed are the foot pound and the erg. Since the erg is a very small quantity, a larger unit called the joule is sometimes used. This unit was named in honor of James Prescott Joule, a famous English physicist, who made important discoveries about work. A *joule* is ten million (10,000,000)

ergs. The relation of gravitational to absolute units of work is expressed as follows:

$$1 \textit{ foot pound} = 32 \textit{ foot poundals}$$

$$1 \textit{ gram centimeter} = 980 \textit{ ergs}$$

EXERCISES. **1.** A weight of 10 pounds is lifted vertically to a height of 10 feet. Find the work done in (a) gravitational units; (b) absolute units.

2. A force of 10 grams acts through a space of 10 cm. Find the work done in (a) gravitational units; (b) absolute units.

3. A horse pulling a load of 1 ton along a smooth road exerts a force of 500 pounds. How much work in foot pounds is done by the horse if the load be hauled a distance of 1 mile?

177. Power. *Power* is the time rate of doing work. It is work divided by time; that is,

$$\textit{power} = \frac{\textit{work}}{\textit{time}} = \frac{w}{t}$$

Units of power. The gravitational unit of power is the *horse power* (h. p.); the absolute unit, the *watt.* When the horse power was first proposed as the unit of power, it was thought that a strong horse was capable of doing about 33,000 foot pounds of work per minute, and this value was originally accepted as the unit of power. A *horse power* then is the expenditure of energy at the rate of 33,000 foot pounds per minute, which is equivalent to 550 foot pounds per second. The actual power of an ordinary horse, however, is only about ¾ h. p. The reason that the unit of horse power, as determined by James Watt, is over size is that when he undertook to determine this unit he used a large Clydesdale, the power capacity of which was greater than that of the ordinary horse. The h. p. of an average man is about ⅐ h. p. It is important to note that while the power of the average man is, for continuous work, far below that of a horse, yet for a short time the man may work at a greater rate than that of a horse power. A person in running upstairs, for instance, may work at a rate considerably

above a horse power. The power of the railroad locomotive varies from 500 to 5000 horse power; that of the engines of our large ocean liners, from 10,000 to 40,000 horse power.

The absolute unit of power is the watt, named in honor of James Watt, the inventor of the steam engine. A *watt* is the expenditure of energy at the rate of ten million (10,000,000 or 10^7) ergs per second. A *kilowatt* (kw) is 1000 watts.

While the horse power is the unit generally used in measuring power, it is not altogether satisfactory for use in the power rating of the gasoline engine, because the foot pound is a gravitational unit. The force of gravity, as we have learned, varies from place to place on the earth's surface. The actual power of a 10 h. p. engine rated at sea level would be somewhat greater than a 10 h. p. engine rated at Denver, Colorado, for example, since the force of gravity at Denver is less than at sea level. To get around this difficulty, our National Bureau of Standards has proposed that the horse power be defined as 746 watts. The units of power, then, may be expressed as follows:

$$\textit{horse power} = \frac{\textit{foot pounds}}{550 \times \textit{seconds}} = \frac{\textit{watts}}{746}$$

$$\textit{watts} = \frac{\textit{ergs}}{10^7 \times \textit{seconds}}$$

EXERCISES. **4.** During the construction of a building, 49,500 lb. of brick were elevated to a height of 20 ft. in 5 minutes. At what rate, in h. p., was the work done?

5. A 500 kilogram weight was lifted to a vertical height of 10 meters in 10 seconds. At what rate was the work done in (a) gram centimeters per second? (b) ergs per second? (c) watts?

6. How many gallons of water can be elevated to a height of 30 ft. by an 8 horse power engine in a quarter of an hour, assuming that a gallon of water weighs 8 lb.?

7. An engine lifts 198 tons of ore per hour from a mine 1000

ft. deep. Find the power of the engine in (a) horse power; (b) kilowatts.

178. Energy. *Energy* is the capacity that a body has for doing work. The steam in an engine possesses energy because in expanding it is capable of doing work. The water in a mill dam, the coiled spring of a watch, and the muscles of the body all possess energy because they have the capacity of doing work.

The units in which energy is measured are the same as the units of work, namely, the foot pound, the foot poundal, the gram centimeter, and the erg.

Energy is of two kinds, potential and kinetic.

179. Potential Energy. *Potential energy* is the energy which a body possesses by virtue of its position, or by virtue of its tendency to change chemically. A body lifted from the floor to the table possesses energy by virtue of its position, because if it were allowed to fall from the table to the floor, it would do work. A lump of coal possesses potential energy because of its affinity for oxygen, that is, its tendency to burn. Gunpowder, likewise, possesses potential energy because of its tendency to explode.

The potential energy of a body due to its position may be measured by the work required to put it in place. For example, a body lies upon the table. Its potential energy with respect to the floor is equal to the work required to lift it from the floor to the table. We then may write

$$potential\ energy = work = force \times space$$
$$P.\ E. = w = Fs$$

EXERCISES. **8.** A piece of metal weighing 10 lb. is lifted from the floor to the table, a distance of 3 ft. Find its potential energy in foot pounds.

9. A mass of 2 kg rests on a shelf 3 m above the floor. What is its potential energy with respect to the floor in (a) gram centimeters? (b) ergs? (c) joules?

10. An automobile weighing 3000 lb. running along a level

stretch of highway comes to a hill the grade (rise) of which is 5 ft. in 100 ft. What potential energy does the car possess with respect to the bottom of the hill when it has traveled a distance of 600 ft. up the incline?

180. Kinetic Energy. *Kinetic energy* is the energy which a body possesses by virtue of its motion. We instinctively avoid bodies that are in rapid motion. Experience has taught us that such bodies possess energy—the greater the motion the greater the energy. Let us consider the case of the bullet and the gun. We have learned from the third law of motion that the momenta of the two are equal. This does not imply, however, that their energies are equal. The energy of the bullet is enormously greater than that of the gun because of its greater velocity. The kinetic energy of a body varies directly as its mass and the *square* of its velocity. If the velocity of a body be doubled, its energy is increased fourfold; if the velocity be trebled, its energy is increased ninefold, and so on. When the mass and velocity of a moving body are given, the kinetic energy is

$$\textit{kinetic energy} = \tfrac{1}{2}\ \textit{mass} \times \textit{velocity squared}$$

$$K.\ E. = \tfrac{1}{2}\, mv^2$$

It is instructive to note how this equation is derived. Since the kinetic energy which a moving body possesses is equal to the work which it can do, and since work = Fs, we have

$$K.\ E. = Fs$$

But in Art. 134 we learned that $v = at$ and $s = \tfrac{1}{2}\, at^2$, and in Art. 142 we learned also that $F = ma$. Now substituting these values of F and s in the equation above, we have

$$K.\ E. = Fs$$

$$= ma \times \tfrac{1}{2}\, at^2 = \tfrac{1}{2}\, ma^2t^2$$

Since $at = v$, then $a^2t^2 = v^2$ and our equation becomes

$$K.\ E. = \tfrac{1}{2}\, mv^2$$

This equation expresses the kinetic energy in *absolute units*. When the mass is in pounds and the velocity in feet per second, the result is given in *foot poundals*. To change foot poundals to foot pounds, divide by 32. When the mass is in grams and the velocity in centimeters per second, the result is in *ergs*. To change ergs to gram centimeters, divide by 980.

Example. A body having a mass of 16 pounds moves with a velocity of 20 feet per second. Find its *K. E.* *Solution*: $K.E. = \frac{1}{2}\, mv^2 = \frac{1}{2} \times 16 \times 400 = 3200$ foot poundals $= 100$ foot pounds.

EXERCISES. **11.** A mass of 64 lb. is moving with a velocity of 10 ft. per second. Find its K. E. in (a) absolute units (foot poundals); (b) gravitational units (foot pounds).

12. A mass of 196 g, moving with a velocity of 10 cm per second, has what K. E. in (a) ergs? (b) gram centimeters?

181. The Conservation of Energy. Potential energy and kinetic energy are so related that when one disappears the other appears. If a stone be thrown upward, it has its maximum kinetic energy at the instant it leaves the hand, because its velocity at that instant is greatest. As it rises, its kinetic energy decreases and its potential energy increases. When it reaches its highest point, its potential energy is a maximum; its kinetic energy, zero. Thus, in ascending there is a transformation from kinetic to potential energy. In descending, a reverse transformation occurs; that is, the potential energy again becomes kinetic energy, reaching a maximum when the velocity is the greatest. Thus it appears that no energy is lost; when one kind disappears another appears.

The law of the conservation of energy states that the energy in the universe is constant in quantity; it cannot be created or destroyed. When work is done, one body loses energy and another gains it. The doing of work, therefore, involves a transference, and often a trans-

formation of energy. The potential energy possessed by a lump of coal is transformed into heat energy in the furnace, then to the steam in the engine, thence to the motion of the wheels, and so on. No energy is destroyed; it is only transformed from one form to another.

Equilibrium and Stability

182. Introductory. In its general dimensions as well as in the enormous size and weight of the structural

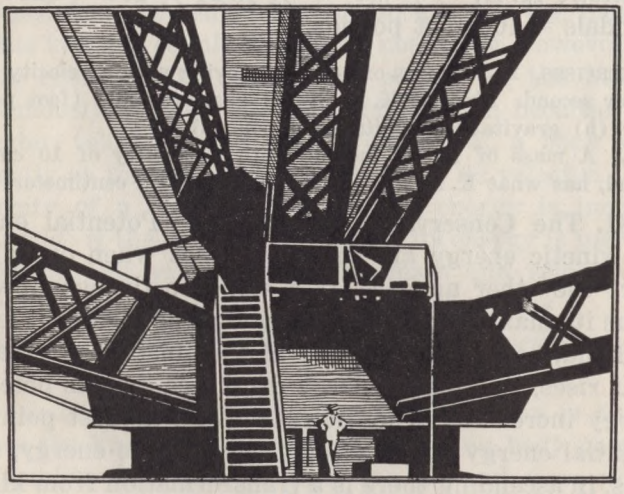


Fig. 147.—One of the supporting piers of the great Quebec bridge, the steel base upon which the truss beams rest weighing 400 tons

members composing it, the Quebec bridge, which spans the St. Lawrence river, surpasses any other structure of the kind ever erected. The one bridge structure in the world that approaches it in magnitude is the famous Firth of Forth bridge in Scotland, the main-channel-span of which is nearly 100 feet shorter than that of the Quebec bridge. Both bridges are of the cantilever

type. The channel-span of the Quebec bridge, measured between centers of towers, is 1800 feet. The design and fabrication of the steel for this structure therefore presented engineering problems for which no precedents existed, and the first attempt to build the bridge resulted in 1907 in a collapse of the structure in which many lives were lost. The present bridge is on the same site as the original structure. The two main piers which support the bridge contain approximately 60,000 cubic yards of masonry and steel. One of these piers, the supporting section of which is shown in Fig. 147, goes to a depth of 60 feet below the bed of the river, and the other to a depth of 80 feet.

One of the shore spans of the Quebec bridge is shown in Fig. 148. A glance at the great steel arm projecting out over the river will at once incite in the mind of the observer the question as to how it is possible to build such a steel structure without support of some kind. The way it is done is to start at the pier as a center and to build the span out equally in opposite directions, that is, toward the shore on one side and toward the center of the river on the other. In other words, the project presents a problem of *equilibrium*. In constructing cantilever bridges of this sort, arms are built out from each shore until they meet, or until they are close enough to receive the central span.

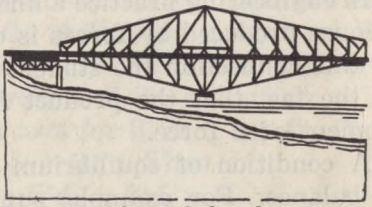


FIG. 148.—One of the shore spans of the Quebec bridge

The problem of equilibrium, as illustrated in the construction of bridges of this type, is one which constantly confronts the structural engineer.

183. Equilibrium. In order to get at the basis of the

problem of equilibrium, it is necessary for us first to get a clear idea of the meaning of the term "moment of a force." The *moment of a force* or *torque*, as it is sometimes called, is the product of a force and its lever arm. Suppose that a force F, Fig. 149, acts downward on the lever arm d which is rigidly clamped at the end O. The force F thus tends to produce rotation about the point O. The force F times its lever arm d, is called the moment of the force, that is

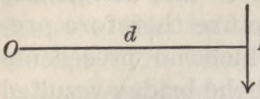


FIG. 149.—Moment of a force, Fd

$$\textit{moment of a force} = \textit{force} \times \textit{lever arm} = Fd$$

in which the force F acts at right angles to the lever arm d.

In engineering practice a knowledge of the moments or torques produced by forces is of the utmost importance. In this connection the student should make special note of the fact that the product Fd is not work; it is the moment of a force.

A condition of equilibrium is essentially a condition of balance. For example, suppose that two masses, A and B, are attached to the ends of a light rigid rod, as shown in Fig. 150. Suppose further that the weight of B is three times as great as that of A, and also that the weight of the rod is negligible as compared with the sum of the weights of A and B. Now the system will manifestly be in equilibrium, that is, in balance, about its axis of support at O when the moment Fd is equal and opposite to the moment $F'd'$. Since F' is three times as great as F, it follows that O must be in such a

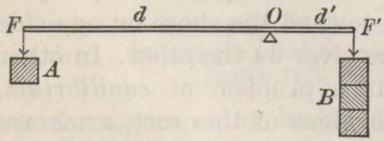


FIG. 150
Condition of equilibrium

position that d is three times as great as d'; that is,

$$Fd = F'd'$$

A body is in equilibrium when the sum of the moments on one side of the axis of support are equal and oppositely directed to the sum of the moments on the other side. When the weight of the lever arm cannot be neglected, we have to take into consideration the weight of the arm, assuming that the weight is a force acting downward at its center of gravity.

Example. Given a rod AB of uniform cross sectional area 12 ft. long, and having a weight W of 10 lb. The rod is placed in a horizontal position on a support which is 4 ft. from the end A. A weight of 20 lb. is suspended from the end B. What weight suspended from the end A will be required to keep the system in equilibrium? *Solution*: Let F be the required force acting at the end A; its lever arm $d_1 = 4$ ft. Acting downward on the other side of the support there are two forces, W and F'. W is a force equal to the weight of the rod acting at its center of gravity; the lever arm for W is $d_2 = 2$ ft. $F' =$ 20 lb. and its lever arm is $d_3 = 8$ ft. Then

$$Fd_1 = Wd_2 + F'd_3$$
$$F \times 4 = 10 \times 2 + 20 \times 8, \text{ or}$$
$$F = 45 \ lb.$$

184. The Problem of Moments Continued. In the preceding article we considered the law of moments as applied to the simple lever. The mathematical expression of this law may be very much more complex than that illustrated by a simple rigid bar. A case in point is that illustrated by Fig. 151. Here we have a bar AB pivoted at A and supported at right angles to the wall AC by means of a cord CB. A force F acts downward on AB tending to produce rotation about the point A. The moment of this force is Fd. An opposing force F' acts on AB along the line BC. The lever arm for this

force (F') is d', a line drawn from A and at *right angles* to CB. The moment in this case is $F'd'$.

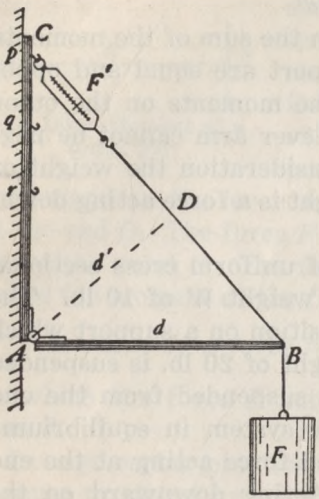


FIG. 151
Moment Fd = Moment $F'd'$

Now since the system is in equilibrium, that is, the bar AB is at rest, the two moments are equal to each other in magnitude and oppositely directed, and we may therefore write,

$$Fd = F'd'$$

In the case of the moment of a force, the lever arm (d or d') is always a straight line measured from the axis of rotation (A in Fig. 151) to the line of direction of the force considered, and at *right angles* to it. Thus d is the lever arm for the force F which acts downward, and d' is the lever arm for the force F' which acts in the direction BC.

Example. A weight of 12 lb. is suspended from the end B of the rod AB, Fig. 151, which is 3 ft. in length. The length of d' is 27 in. What is the tension (pull) in the string CB, as recorded by the spring balance at F'?

Solution:

$$Fd = F'd'$$

$$12 \times 36 = F' \times 27, \textit{ hence } F' = 16 \textit{ lb.}$$

185. Three States of Equilibrium. The three states of equilibrium which a body may possess are stable, unstable, and neutral, illustrated by a cone as shown in Fig. 152. A body is said to be in *stable equilibrium* if a turning motion about any point in its base, such as would occur in tipping it over, tends to raise its center

of gravity (Art. 8). A book lying on the table is in stable equilibrium because its center of gravity rises if it be turned upon its edge. A cone resting upon its base is a good illustration of stable equilibrium.

A body is in *unstable equilibrium* if a turning motion tends to lower its center of gravity as in the case of a cone standing on its vertex.

A body is in *neutral equilibrium* if a turning motion neither raises nor lowers its center of gravity. A spher-

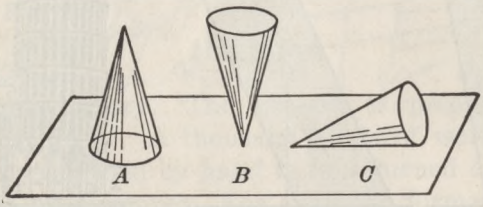


FIG. 152.—Three states of equilibrium

ical ball is an excellent illustration of a body in a state of neutral equilibrium. The cone *C*, Fig. 152, may be considered to be in neutral equilibrium, so far as a rolling motion is concerned.

186. Stability of a Body. The *line of direction* is a straight line passing through the center of gravity of a body and the center of the earth. Whether a body stands or falls depends upon the position of its line of direction with respect to its base. If this line falls within the base, *I*, Fig. 153, the body will stand; if it falls without the base *II*, the body will fall. There are in Europe a number of "leaning towers" which illustrate in a very striking manner the relation of the stability of a body to the position of its line of direction. One of the most famous of the leaning towers is that of the bell tower of the Cathedral of Pisa, Italy, Fig. 154. The tower is 179 feet in height

and its top leans from the vertical a distance of more than 16 feet.

The amount of stability which a body possesses is determined by the work required to overturn it, that is,

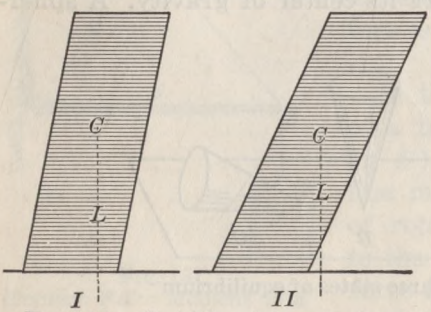


FIG. 153.—Stability determined by line of direction L



FIG. 154
The Leaning Tower of Pisa

the amount of work necessary to lift its center of gravity through a vertical distance ab, Fig. 155. There are two ways by which the stability of a body may be increased, namely, (a) by broadening the base, (b) by lowering the center of gravity. In A, B, and C of Fig. 155 we have three bodies of the same shape and size. In C the center of gravity is lower than in A and B, due to the loading of one end of the vessel with some heavy material.

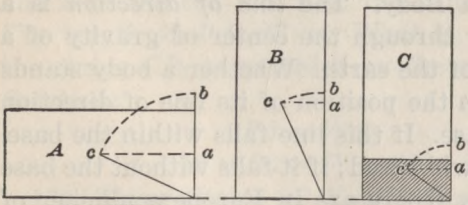


FIG. 155.—Condition of stability

EXERCISES. **13.** Explain (a) why *A*, Fig. 155, is more stable than *B*; (b) why *C* is more stable than *B*.

14. Which is the more stable, a load of hay or a load of coal, each weighing the same? Why?

15. A person standing with his heels and back against a wall cannot pick up anything from the floor in front of him without falling. Why?

16. Why does a person carrying a heavy bucket of water throw out the free arm, as shown in Fig. 156?



FIG. 156

MACHINES

187. Introductory. The present is preeminently an age of machinery. A thousand kinds of work that was formerly performed by hand is now turned out in great quantities by machines. For example, formerly a tailor cut out a single piece of cloth with handshears; now a machine operator cuts at a single stroke fifty pieces. In the old days a shoemaker could turn out a single pair of shoes in a day; now wonderfully complex machines turn out several hundred times as much work in a day as was done by the old-fashioned shoemaker. Machines have been designed and are now in use which are capable of soling 600 pairs of shoes in a single day.

It must not be understood, however, that the term "machine" has reference only to such complicated pieces of apparatus as mentioned in the preceding paragraphs. A jackknife is a machine, as is also a lead pencil, and a sewing needle. In a scientific sense a *machine* is a device for transferring or transforming energy: A steam engine together with the boiler, is a machine for transforming the potential energy of coal into the kinetic energy of mechanical motion. A dynamo is a machine which may be used for transforming the energy of me-

chanical motion into the energy of an electric current. A hammer, a pair of scissors, or a hand spike are all types of machines for transferring energy from one point to another.

All machines, however complicated, are made up of one or more of the six so-called *mechanical powers*, or simple machines. These are the *lever, wheel and axle, pulley, inclined plane, wedge, and screw.*

In elementary physics it is customary, in solving problems relating to simple machines, to consider the friction factor as negligible and to assume, usually, that the parts of a machine are rigid and without weight. When these factors (friction and weight of the moving parts) are taken into account, problems relating to machines are sometimes very complicated and require for solution the principles of advanced mathematics.

188. Law of Machines. In accordance with the law of the conservation of energy, the work done by a force acting on an ideal frictionless machine must equal the work done by the machine. This principle as applied in mechanics is expressed by the general law of machines: *The force, multiplied by the distance through which it acts, is equal to the resistance overcome, multiplied by the distance through which it acts.* This statement of the law, of course, takes no account of friction.

The general law of machines may be demonstrated as follows: Weight the end of a rod so that it will balance about some point, for example, one-third of the distance from one end, Fig. 157. This constitutes a simple machine. To one end of the lever thus

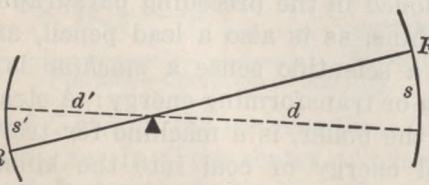


FIG. 157.—Illustrating the law of machines

formed suspend a mass of 1 unit of force, F, and to the other end 2 units of resistance, R. The system is in equilibrium. If F be moved through a distance s, R will move through a distance s'. The force applied to a machine is usually called the *effort*. In accordance with the law of machines we may therefore write

$$effort \times space = resistance \times space$$
$$Fs = Rs'$$

Since the two figures bounded by the arcs s *and* s', Fig. 157, are similar, we may substitute d and d' for s and s' and our equation becomes one of moments, that is

$$Fd = Rd'$$

189. Mechanical Advantage. Several advantages may be derived from the use of a machine. For example, first, we may apply a force in the most advantageous direction, as in the lifting of a weight by passing a rope over a pulley. Second, we may gain in speed at the expense of force, as exemplified in the gearing of a bicycle. And third, we may use a small force to overcome a large resistance, as in the case of lifting a heavy weight by means of a lever. In general when we speak of the mechanical advantage of a machine, we use the term in the third sense, as stated above.

Mechanical advantage is the ratio of the resistance to the effort, that is,

$$mechanical\ advantage = \frac{resistance\ overcome}{force\ applied} = \frac{R}{F}$$

In determining the mechanical advantage of a machine, two things should be noted: first, in computing the resistance R, *friction is not taken into account,* and second, it is frequently desirable to express the ratio of R to F in terms of certain parts of the machine, as will be explained later.

190. Mechanical Efficiency. The *efficiency* of a machine is the ratio of the useful work gotten out of the machine to the total work put into it. In equational form

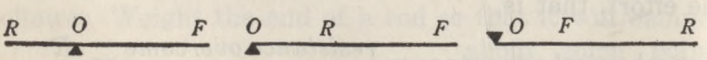
$$\textit{efficiency} = \frac{\textit{work done by the machine}}{\textit{work done on the machine}} = \frac{\textit{work out}}{\textit{work in}}$$

For example, if a machine delivers 60 foot pounds of useful work for every 100 foot pounds put into it, its efficiency is 60/100 = 60 per cent. This means that 40 per cent of the energy is lost through friction or through other causes.

A perfect machine, that is, one operating without friction, would have an efficiency of 100 per cent. Since it is impossible to make such a machine, the work done by a machine is always less than the work put into it. The efficiency of machines varies greatly, depending on the amount of friction encountered. In the case of a simple lever the efficiency may be nearly 100 per cent while for the screw it may be as low as 25 per cent. The difference between the work done on a machine and the useful work delivered by it represents the frictional losses. In actual practice there is always some waste energy, hence we may write

$$\textit{work in} = \textit{work out} + \textit{frictional losses}$$

191. The Lever. A *lever* is a rigid bar capable of moving about a fixed support called a fulcrum. There are



First class Second class Third class

FIG. 158.—Three classes of levers

three classes of levers, as shown in Fig. 158, in which F is the force applied, R the resistance overcome, and O the fulcrum. In each case $OF = d$ and $OR = d'$, in which d is called the force or effort arm and d' the resistance arm. From Fig. 158 we have

first class : *fulcrum between F and R*
second class : *fulcrum at end, R being between O and F*
third class : *fulcrum at end, F between O and R*

The following points should be noted with respect to the lever: (a) The force arm d is measured from the point of application of the force F to the fulcrum O; the resistance arm d', from the resistance R to the fulcrum. (b) In the case of a bent lever, the arms d and d' are measured from the fulcrum to the lines of direction F and R, and at right angles to the same, Fig. 159. (c) Since for levers of the same length the force arm d is the greatest for the second class, it follows that for a given length, a lever of the second class offers the greatest mechanical advantage. (d) A lever of the third class gives an advantage in speed at the expense of the force applied.

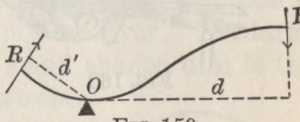


FIG. 159

The law of the lever is derived from the general law of machines, namely

$$effort \times force\ arm = resistance \times resistance\ arm$$
$$Fd = Rd'$$

from which equation we may write

$$mechanical\ advantage = \frac{R}{F} = \frac{d}{d'}$$

Example. Given a light rigid bar 6 ft. in length which is used successively as a lever of the first class, second class, and third class, to find the mechanical advantage under the following conditions: First class, fulcrum 2 ft. from R; second class, R 2 ft. from the fulcrum, which is at the end of the bar; third class, F 2 ft. from fulcrum. *Solution, first class*: Since the fulcrum is 2 ft. from R, $d = 4$ ft. and $d' = 2$ ft. Mechanical advantage $= d/d' = 4/2 = 2$. *Solution, second class*: In this case $d = 6$

and $d' = 2$, hence the mechanical advantage $= d/d' = 6/2 = 3$. *Solution, third class*: Here $d = 2$ and $d' = 6$, and therefore the mechanical advantage $= d/d' = 2/6 = 1/3$. In this case speed is gained at the expense of the resistance overcome.

EXERCISES. **17.** Classify the levers of Figs. 160, 161, 162, 163, 164, and 165 with respect to class.

18. If a man exerts a force of 100 lb. on a lever of the second class, at a point 6 ft. from the fulcrum F, what weight can be lifted, provided the resistance arm d' is 18 in. in length?



FIG. 160



FIG. 161

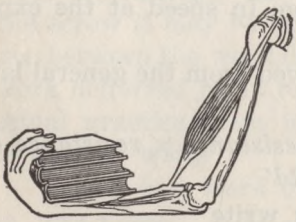


FIG. 162



FIG. 163



FIG. 164

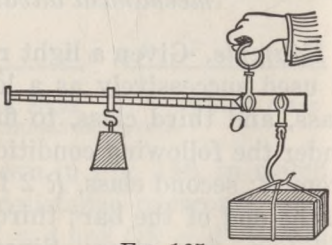


FIG. 165

19. If a weight of 300 lb. acts downward at the center of gravity of the wheelbarrow, Fig. 164, which is 1 ft. from

the axis of the wheel (fulcrum), what upward force must the man exert, providing the distance from the center of gravity to his hands is 3 ft.?

20. How many levers is the baggage-man shown in Fig. 166 using? To what class do these levers belong, assuming that the center of mass of the box is at its geometrical center?



FIG. 166

192. The Wheel and the Axle. The wheel and axle is an application of the principle of the lever. The large disc, Fig. 167, is the wheel; the small shaded disc is the axle. The force F is applied to a rope passing over the wheel; the resistance or weight R is lifted by means of a rope around the axle. The center of the system O is the fulcrum, $FO = r$ is the force arm, and $OR = r'$ is the resistance arm, in which r and r' are the radii of the wheel and axle, respectively. Since the radii of two wheels are proportional to their circumference (c and c'), these values may be substituted for d and d' in the equation $Fd = Rd'$. The law of machines as applied to the wheel and axle then becomes

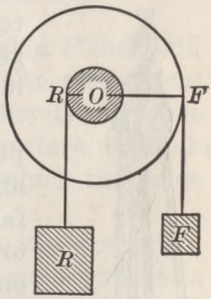


FIG. 167
Wheel and axle

$$\textit{effort} \times \textit{radius large wheel} = \textit{resistance} \times \textit{radius small wheel}$$

$$Fr = Rr', \text{ or}$$

$$Fc = Rc'$$

in which c is the circumference of the large wheel and c' that of the small wheel.

EXERCISES. **21.** What force applied at F, Fig. 167, will support

a weight of 100 lb., if the radius of the large wheel is 2 ft. and that of the small wheel (axle) 6 in.?

22. Suppose that for a given wheel and axle, similar to that of Fig. 167, the circumference of the wheel is 5 ft. Find the circumference of an axle such that a force of 20 lb. will support a weight of 100 lb.

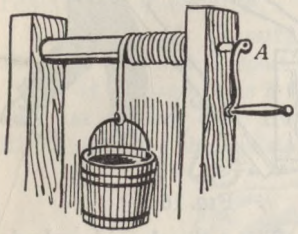


FIG. 168.—Windlass

23. If the length of the lever arm A of the windlass, Fig. 168, is 1.5 ft., and the circumference of the axle is 1.5 ft., what weight will be supported by a force of 10 lb.? When the handle makes one revolution, through what distance will the weight attached to the rope move?

193. The Pulley. A *pulley* is a wheel turning about an axis in a frame or block. A set of blocks containing one or more pulleys each, together with the attached rope, is called a block and tackle, Fig. 169.

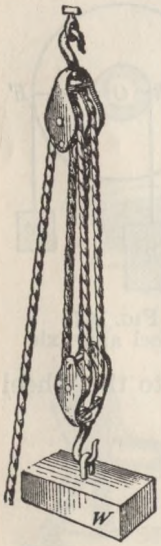


FIG. 169
Block and tackle

With respect to the point of support, pulleys are of two kinds, fixed and movable. A *fixed pulley* is one that is fixed or fastened to a stationary support, as a beam or a wall, Fig. 170 *A*. A *movable pulley* is one that is attached to the weight or object to be moved. Fig 170 *B* illustrates a system of one fixed and one movable pulley.

When F moves downward through a distance s, R moves up through a distance s'; hence the law of the pulley is $Fs = Rs'$ and

$$\textit{mechanical advantage} = \frac{R}{F} = \frac{s}{s'}$$

As applied to the pulley, the law of machines, $Fs = Rs'$, may be experimentally verified for a number of typical cases by

means of two or three pulleys and a spring balance, arranged as in the following figures:

(a) Fig 170 *A*. In this case when F moves down one unit of distance, R moves up one unit, that is $s = s'$, and

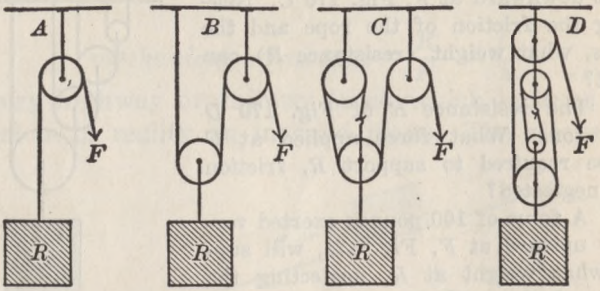


FIG. 170.—Mechanical advantage of the pulley

hence $F = R$. A force of 10 pounds applied at F then will balance a resistance at R of 10 pounds. The mechanical advantage in the case of a single fixed pulley is therefore $R/F = 1$. The only advantage gained, therefore, by such a pulley is a change in the direction in which the force acts.

(b) Fig. 170 *B*. Since we have here two ropes attached to the movable pulley in this case, F will move two units of space for R one unit; hence 1 unit of force at F will support 2 units of resistance at R. A force of 10 pounds at F will therefore exert a force of 20 pounds at R. The mechanical advantage in this case is $R/F = 2$.

(c) Fig. 170 *C*. Here three ropes are attached to the movable pulley. When F moves 3 units of distance, R will move 1 unit; therefore 10 pounds of force at F will exert 30 pounds at R. The mechanical advantage here is $R/F = 3$.

(d) Fig. 170 *D*. Four ropes are pulling (due to the force F) upon the movable pulley. When F moves through

4 units of space, R moves through 1 unit; a force of 10 pounds applied at F will therefore overcome a resistance R of 40 pounds. The mechanical advantage is $R/F = 4$.

EXERCISES. **24.** A man weighing 160 pounds exerts the force of his entire weight downward at F, Fig. 170 *C*. Neglecting the friction of the rope and the pulleys, what weight (resistance R) can he lift?

25. The resistance R of Fig. 170 *D* is 2.5 tons. What force applied at F will be required to support R, friction being neglected?

26. A force of 100 pounds exerted vertically upward at F, Fig. 171, will support what weight at R, neglecting the weight of the pulleys and the friction factor?

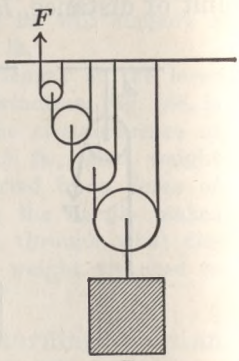


FIG. 171

194. The Inclined Plane. An *inclined plane* is any plane surface inclined at an angle with the horizontal; it is a device employed for lifting heavy bodies against gravity by sliding or rolling them up the incline, Fig. 172.


FIG. 172

Let us consider the incline plane shown in outline in Fig. 173. A force F applied parallel to the incline is employed to push the heavy weight R up the incline frame A and C, thus lifting the block vertically against gravity through the distance BC. Neglecting friction, the product of the force F and the distance AC through which it acts is equal to the product of the weight of the block R and the vertical distance BC through which it is lifted. If we let l equal the length

of the incline AC and h the vertical distance BC, we may then write the equation for the inclined plane

$$effort \times length\ of\ incline = weight \times vertical\ height$$

$$Fl = Rh$$

When the force F is applied parallel to the incline we have

$$mechanical\ advantage = \frac{R}{F} = \frac{l}{h}$$

Every highway or railway track which inclines from a level is in reality an inclined plane. The steeper the

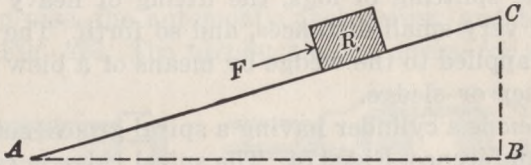


FIG. 173.—Inclined plane

grade the greater the force required to move the car up the slope. By *grade* we mean the vertical raise in feet per 100 feet of base line, measured along AB. For example, if a road rises 2 feet for every 100 feet of base line, its grade is $2/100 = 2$ per cent. Railroad grades rarely exceed 2 per cent; highway grades may in some instances be as great as 8 to 10 per cent. Since grade is the ratio of the vertical height BC, Fig. 173, to the base line AB, we may write:

$$grade = \frac{BC}{AB} = \frac{h}{b}$$

Example. The grade of a certain hill on a highway is 5 per cent. How heavy a load can be pulled up this hill by a horse that is capable of exerting a pull of 200 lb.? *Solution*: Since the grade is 5 per cent, $h = 5$ and $l = 100$. From the law of the inclined plane we then have

$$Fl = Rh$$

$$200 \times 100 = R \times 5, \text{ hence } R = 4000\ lb.$$

EXERCISE 27. (a) Neglecting friction, what force applied parallel to the incline will be required to support a 200 lb. weight on a smooth plane, the incline of which is 16 ft., and the vertical height 4 ft.? (b) What is the mechanical advantage of this plane?

195. The Wedge and Screw. A *wedge* is a modified form of an inclined plane. It may be considered as being made up of two planes placed base to base. The blade of a pocket knife, the end of a nail, the point of a needle are all wedge-shaped instruments. The wedge is usually used to overcome great resistance through small spaces, as in the splitting of logs, the lifting of heavy weights through very small distances, and so forth. The force is usually applied to the wedge by means of a blow as from a hammer or sledge.

A *screw* is a cylinder having a spiral groove cut around its circumference, Fig. 174. The spiral ridge is called the thread and the distance between two consecutive threads the *pitch*. The mechanical advantage of the screw is derived from a combination of the principles of the lever and the inclined plane.

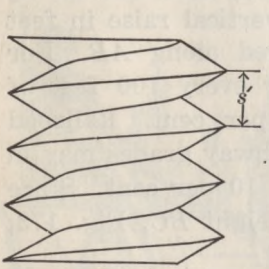


FIG. 174
Pitch of screw = s'

When the screw is rotated once, it moves forward in the direction of its axis through a distance s' equal to the pitch. The screw usually is turned by means of a lever or wheel. The force is applied to the end of the lever, or at the circumference of the wheel. While the force moves once around the circumference of the wheel the resistance R acts through a distance equal to the space s' between two consecutive threads (the pitch of the screw). The distance through which the force F is exerted during one revolution is $s = 2\pi r$, where r is the radius of the wheel. The work done therefore equals $F \times 2\pi r$, and the work

done in overcoming the resistance is Rs'. We may therefore write

$$Fs = Rs'$$
$$F \times 2\pi r = Rs'$$

from which

$$\textit{mechanical advantage} = \frac{R}{F} = \frac{2\pi r}{s'}$$

The mechanical advantage of the screw is relatively high since $2\pi r$ is usually great as compared with s'.

Applications of the screw are manifold. For example, we have the common screw, the bolt and nut, Fig. 175, the shop vise, the automobile jack screw, and the turnbuckle, Fig. 176. The turnbuckle is a device for bringing

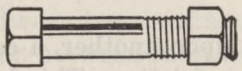


FIG. 175.—Bolt

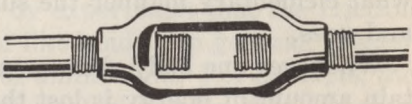


FIG. 176.—Turnbuckle

the ends of two rods together. It is a double headed bolt. Since there are two threaded ends of the rods, each turn of the nut brings the rods together by a distance equal to 2 times the pitch of the screw.

EXERCISES. **28.** The distance from A to B, Fig. 177, is one inch. (a) How many threads to the inch on this screw? (b) What is the pitch s'? Express your answer in the fraction of an inch.

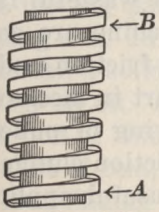


FIG. 177

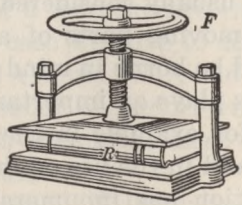


FIG. 178

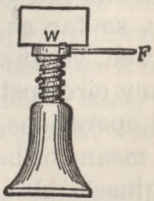


FIG. 179

29. If a force of 20 lb. applied to the circumference of the wheel of the book press, Fig. 178, acts through a distance of 3 ft.

in making one complete revolution, and thus causing the screw to move downward ¼ in., what will be the resistance overcome at R?

30. The lever WF of the jack screw, Fig. 179, is 1 meter in length. If one revolution of the effort causes the screw to move upward 1 cm, what resistance will be overcome by a force of 10 kg?

31. The turnbuckle shown in Fig. 176 has 16 threads to the inch. The ends of the rods to which it is threaded are 1 in. apart. How far apart are they after 4 complete turns?

FRICTION

196. Introductory. Thus far in our discussion of machines we have neglected to take into account the factor of friction; in other words, we have been dealing with ideal simple machines. We shall now consider in a somewhat elementary manner the subject of friction, its laws and uses.

Whenever one body slides or rubs upon another, a certain amount of energy is lost through friction. This loss of energy appears mainly as heat. *Friction* is the resisting force which is encountered when one body slides or rolls upon another.

Friction may occur between solids and solids; between solids and liquids, as in the case of the passage of a boat through water; and between solids and gases, as exemplified by the friction encountered by the passage of an airplane through the air.

While friction is usually considered as a wasteful factor, so far as the moving parts of a machine are concerned, yet it should be borne in mind that friction under many circumstances plays an important part in mechanical operations, as for example in the driving of pulleys by means of belts, and in the use of the friction clutch in engines. Also, friction has innumerable useful applications in the affairs of everyday life. Nails and screws are held in place by friction, we are enabled to walk upon smooth floors and other smooth bodies because of

friction. Other illustrations might be cited in great numbers.

197. Fundamental Facts Relative to Friction. (a) Friction between two bodies, A and B, as for example a brick lying upon a smooth plank, is mainly due to the inequalities which occur upon the surfaces of the two bodies. It is evident that the force required to push or pull A over the surface of B will depend to a large extent on how closely the irregularities of the surface of A fit into those of B. It is clear, also, that the greater the force pressing A and B together (normal force) the more intimate will be the contact between them, and hence the greater will be the frictional force.

(b) It has been found by experiment that the friction between two substances of like kind is, in general, greater than between substances of unlike kind. For example, the friction between steel and steel is greater than between steel and brass. It is for this reason that steel shafts are frequently set to rotate in brass bearings. For the same reason also, "babbitt" bearings are employed, the friction between steel and babbitt being relatively small. Babbitt metal is a soft alloy, extensively used for the linings of axle bearings, in order to reduce friction; it is an alloy consisting of copper, tin, and antimony.

(c) A common method of reducing friction between the moving parts of a machine involves the use of lubricating oils. In this case we have to deal with the friction between a solid and a liquid (oil), and also between a film of oil on oil. A number of factors enter into a study of friction of this sort, among the most important of which is that of temperature.

198. Sliding Friction. Friction may be classified as *sliding friction* and *rolling friction*. Sliding friction, sometimes called kinetic friction, is the resisting force encountered when one body slides with uniform speed

over another body. In this case the pulling force is just equal and opposite to the frictional force. The following facts have been determined experimentally with reference to kinetic friction.

(a) Kinetic friction is directly proportional to the force with which the surfaces are pressed together. Thus, if one brick drawn over the table offers a given resistance, two bricks, one placed upon the other, will offer twice as much resistance (friction).

(b) Kinetic friction is independent of the area of the moving body, provided the moving body does not cut into the surface upon which it slides. To prove this attach

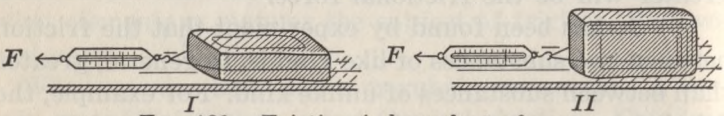


FIG. 180.—Friction independent of area

a spring balance to a brick, the broad surface of which is in contact with the table, as shown in Fig. 180, *I*. By means of the spring balance, draw the brick over the surface of the table with a uniform motion and note the reading of the balance. Now place the brick with its narrow face in contact with the table, *II*, and draw it along as before. The reading of the balance will be the same as in the first case, thus showing that the friction is independent of the area of the surface in contact.

(c) Through a considerable range kinetic friction is independent of the speed of the moving body. At very high speed, however, the friction is considerably less than for low speed.

199. Coefficient of Kinetic Friction. In designing machinery in which moving parts are employed, the engineer has to take into consideration a factor known as the coefficient of friction. Friction coefficients of all ordinary

materials are given in tabular form in engineering handbooks.

The *coefficient of kinetic friction* for the substances A and B is the ratio of the force F required to move A over B at a constant speed, to the normal force N pressing the two surfaces together; that is,

$$\textit{coefficient of kinetic friction} = \frac{\textit{moving force}}{\textit{normal force}} = \frac{F}{N}$$

Suppose that we wish to find the coefficient of kinetic friction between iron and leather, as exemplified by the friction between an iron pulley and a leather belt. Place the leather belt on a table in a horizontal position. Place upon the leather a rectangular block of iron, having slightly rounded edges. Suppose that the weight of the iron is 1050 grams (gravitational units). Since the leather surface is horizontal, the normal force N pressing the iron and leather together is equal to the weight of the iron, hence $N = 1050$ grams of force. Now suppose that a pulling force F of 588 grams is required to keep the iron sliding over the leather with constant speed. Then we have in this case

$$\textit{coefficient of friction} = \frac{F}{N} = \frac{588}{1050} = 0.56$$

The coefficient of kinetic friction for dry oak on oak, fibres parallel, is 0.48; steel on steel, not lubricated, 0.21; cast iron on brass, dry, 0.19; iron on leather, dry, 0.56; iron on leather, oily, 0.15.

200. Rolling Friction. *Rolling friction* is the friction encountered by a wheel rolling without slipping upon a surface. When a heavy body such as a cylindrical disc or wheel rolls over a given surface there appears a slight elevation or hump in front of the advancing disc. This elevation is very much exaggerated in Fig. 181, for the purpose of illustration. At every instant during its forward

motion the wheel is in the act of ascending the elevation. Rolling friction, therefore, is measured by the force required to keep the wheel from rolling back. In some cases, as for example, the case of an automobile running over a hard pavement, the deformation occurs in the wheels.

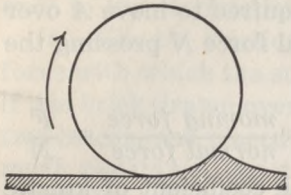


FIG. 181.—Rolling friction

The following facts are known with reference to rolling friction:

(a) Rolling friction is very much less than sliding friction. It is for this reason that it is much easier to draw a given load in a wheeled vehicle over a pavement

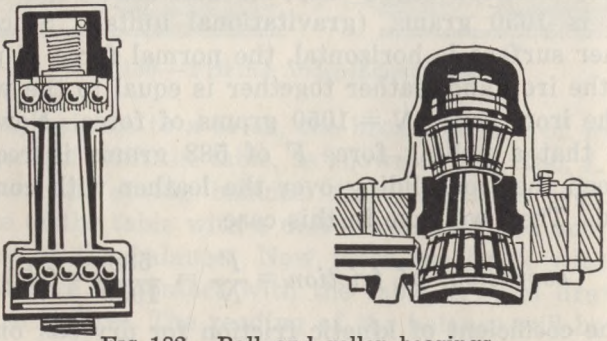


FIG. 182.—Ball and roller bearings

than it is to draw the same load in a sled. For a similar reason, also, ball and roller bearings, Fig. 182, are substituted for bearings in which sliding friction occurs.

(b) Rolling friction is inversely proportional to the radius of the wheel. This explains the advantage of using as large a wheel as practicable for a given surface and load.

(c) Rolling friction is proportional to the weight of the rolling system.

Applications of the Simple Mechanical Powers

201. Train of Gear Wheels. In Fig. 183 there is shown a train of three *gear wheels.* This is an application of the wheel and axle principle, giving a relatively great mechanical advantage. The small toothed wheel which is attached to the axle and which connects (meshes) with the circumference of a larger wheel is called a *pinion.*

In order to determine the mechanical advantage of such a train of wheels, let us begin with the large left hand wheel. There are 40 teeth, or cogs as they are called, in the circumference of this wheel and 10 in the pinion of the middle wheel. The ratio is 4 to 1. One revolution of the large wheel will therefore produce 4 revolutions of the middle wheel. The number of cogs in the circumference of the middle wheel is 32 and the number in the pinion of the right hand wheel is 8. The ratio between these two is also 4 to 1. Four revolutions of the middle wheel will thus produce 16 revolutions of the right hand wheel. That is, one revolution of the left hand wheel will produce 16 revolutions of the right hand wheel. The mechanical advantage, then, so far as the number of cogs on each wheel is concerned, is 16. But we have also to take into account the ratio of the circumference of the left hand wheel, around which is wound the rope to which the force F is attached, to the circumference of the axle of the large wheel, to which R is attached. Let us suppose that the ratio of these circumferences is 3 to 1. The total mechanical advantage of the system then is

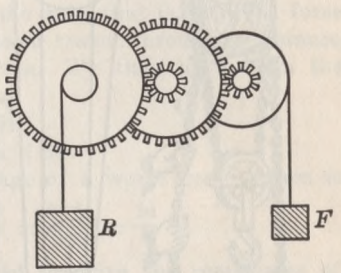


FIG. 183

$$M.\ A. = \frac{no.\ cogs\ wheel\ 1}{no.\ cogs\ pinion\ 2} \times \frac{no.\ cogs\ wheel\ 2}{no.\ cogs\ pinion\ 3} \times \frac{circum.\ wheel\ 3}{circum.\ axle\ 1}$$

This means that if the force F moves down a distance of 48 inches, say, the resisting force R will move up a distance of 1 inch. Or, to look at it the other way around, when R moves through a space of one unit, F will move through a space of 48 units. A train of gear wheels thus gives us a means of multiplying the force or of increasing or decreasing the speed.

202. Differential Pulley. The *differential pulley*, which is a combination of the wheel and axle and the pulley, Fig. 184, is used for lifting heavy weights by hand power. The upper block contains two rigidly connected pulleys of different radii, r and r'; the lower block contains one pulley. A closed chain is used to connect the two blocks.

Let us see if we can work out an expression for the mechanical advantage of this system. Suppose that the operator takes hold of the chain, as shown, and pulls down a distance sufficient to produce one revolution of the connected pulleys in the upper block. The length of chain wound *off* the smaller pulley is $2\pi r'$ and the length wound *on* the larger pulley is $2\pi r$. One movement is down and the other is up, therefore the actual shortening of the chain is $2\pi r - 2\pi r' = 2\pi(r - r')$. Since the lower pulley is supported by two strands of chain, the distance s' that the weight R is lifted is $\frac{1}{2} \times 2\pi(r - r') = \pi(r - r')$. The distance s through which the force F moves is $2\pi r$. Now from the general law of machines, $Fs = Rs'$, we have for the differential pulley

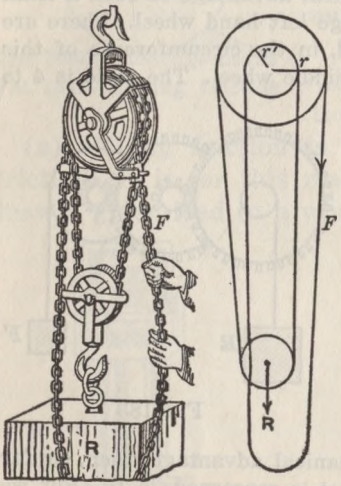


FIG. 184.—Differential pulley

$$F \times 2\pi r = R \times \pi(r - r')$$
$$Fr = \tfrac{1}{2} R(r - r')$$

from which we have

$$\text{mechanical advantage} = \frac{R}{F} = \frac{2r}{r - r'}$$

Since the difference $r - r'$ may be made as small as we choose, it is obvious that as r approaches r' in length, the mechanical advantage correspondingly increases.

203. Worm Gear. A *worm gear*, Fig. 185, is a combination of the lever, screw, and toothed wheel. This device, therefore, gives a very great mechanical advantage, and is frequently used on the

rear drive of auto-trucks. It consists of a screw which is attached to a shaft and which is tangent to a cog wheel. One complete turn of the shaft causes the cog wheel to rotate through a distance equal to the space between two successive cogs of the wheel. If the wheel has n cogs, the shaft then must make n turns in order to produce one revolution of the wheel. Suppose that a force F is applied to the end of the lever of length l and the resistance R is attached to a rope wound around the axle of the wheel, the radius of the axle being r. Then to move R through a distance equal to the circumference of the axle, that is, $2\pi r$, the force F will have to travel through a distance equal to $2\pi ln$. We then have from the general law

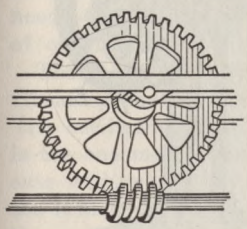


FIG. 185

$$Fs = Rs'$$

$$F \times 2\pi ln = R \times 2\pi r$$

from which the mechanical advantage of a worm gear system is

$$\textit{mechanical advantage} = \frac{R}{F} = \frac{2\pi ln}{2\pi r} = \frac{ln}{r}$$

The worm gear is commonly used to drive the rear axles of motor trucks and also to reduce the speed in such devices as the speed counter.

204. The Sewing Machine. The principal mechanical powers employed in the operation of the sewing machine are the lever, the inclined plane, and the wheel and axle. The machine is run by foot power applied to the treadle or by power supplied directly to the fly-wheel by an electric motor. Running lengthwise along the interior of the arm of the machine there is a crank shaft which is connected to the fly-wheel, and which, in the type of machine shown in Fig. 186, is provided with three crank systems A, B, and C. Crank A operates the pressure-foot rod and also the needle-bar, to the lower end of which is fastened the needle N. Crank B operates a vertical lever of the first class which drives the feed-bar. When the needle-bar is up, the pressure-foot is slightly lifted, at which instant the cloth is pushed forward by the feeding device through the distance of one stitch. When the needle descends the pressure-foot is down, holding the cloth firmly in place. Crank C operates the shuttle-bar.

The most complicated mechanism of the sewing machine is that

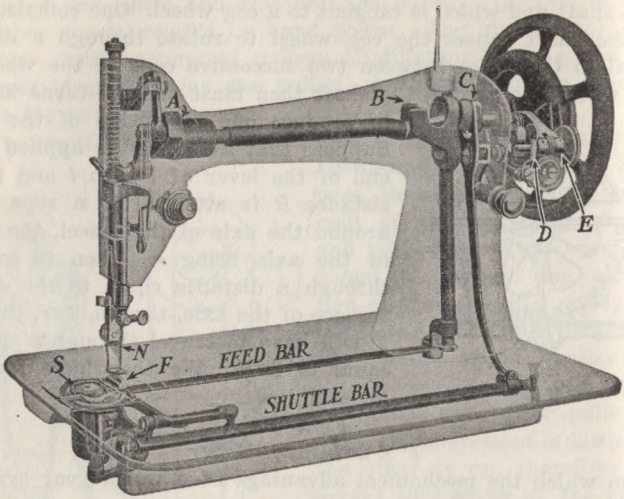


FIG. 186

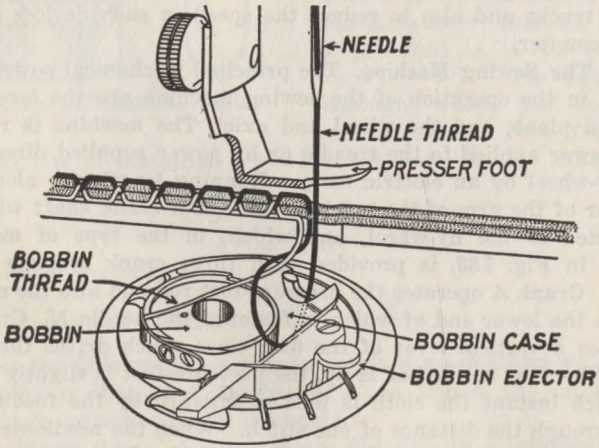


FIG. 187
Mechanical layout of a sewing machine, showing the principal working parts

which includes the shuttle or threading device, Fig. 187. There are two thread-conveying systems, one carrying the thread from the spool down through the needle *N*, and the other operating the bobbin thread which is wound on a small spool or bobbin in the shuttle. When the needle *N* is thrust downward through the cloth, it is necessary that the bobbin thread be passed between the needle-thread and the needle, so as to lock the stitch. This looping of one thread through the other is accomplished by the shuttle.

There are three types of shuttles in common use, namely, the vibrating shuttle, the oscillating shuttle, and the rotating shuttle. The oscillating shuttle, details of which are shown in Fig. 187, is the one most generally used today, because of the comparative ease with which the bobbin may be inserted and removed. When the bobbin is to be rewound with thread, it may be removed from the oscillating shuttle by simply pressing the bobbin ejector. The bobbin is rewound by attaching it to the shaft at the small wheel *E*. The thread is evenly wound on the bobbin shaft by the forward and backward motion of the small lever *D*. The interlocking of the two threads, that is, the needle-thread and the bobbin-thread, is accomplished by the forward and backward motion of the shuttle. During each oscillation of the shuttle, the bobbin-thread is looped through the needle-thread as shown. When this looping is accomplished, the threads are drawn tight by the upward motion of the needle bar. The length of the stitch is regulated by adjusting the length of the stroke of the feed-bar.

205. The Typewriter. The essential working parts of a typewriter include a system of levers of the first, second, and third classes. The key-bar lever system constitutes the heart of the typewriter because motion imparted to the key by the finger of the operator produces the typed page, releases the carriage after each impact of the type, changes the placement of the ribbon after each stroke, and causes the shift from lower case letters to upper case, that is, from small letters to capitals, as well as the shift to the various characters which are usually found around the margin of the key board. While the typewriter is extremely complicated, due to the employment of a great number of lever systems, its fundamental action is not difficult to understand. In Fig. 188 there is shown a single type-bar lever system. Different makes of typewriters employ various combinations of levers. The fundamental principle involved, however, is essentially the same in all cases. For example, the type-bar shown in Fig. 188 is a lever

of the third class, while in some machines it operates as a lever of the first class.

On the downward stroke of the key K, the key-bar acts as a lever of the second class, moving about the fulcrum F. When the

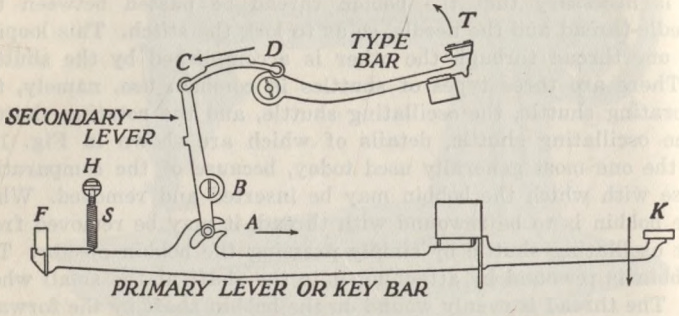


FIG. 188.—The type-bar lever system of a typewriter

pressure of the finger on K is removed, the spring S pulls the lever system back into position, in which case KF acts as a lever of the third class.

The secondary lever ABC is of the first class, moving about the fulcrum B. The lower end of this lever carries a small wheel which moves along the curved groove in the key-bar at A. The object of this grooved path is to give a free, forward accelerated motion to the type-bar as it strikes the paper. Motion is transmitted from the secondary lever to the type-bar by means of the connecting link CD.

The type-bar is a lever of the third class, the force arm of which is very short. As a consequence, the type T moves through a relatively large distance for a small motion of the key K. The travel of the type is often fifteen times as great as that of the space moved by the key.

The trains of levers for each key are arranged so that the ends of the type-bars strike at the same point. The lever bearing the key also works other levers which release the escapement in a fashion similar to the action of a pendulum of a clock. After each stroke of a key this escapement allows the coiled spring, acting through a wheel and axle, to move the carriage containing the paper the distance necessary for printing the next letter. The

space bar, which is a key not connected to any type-bar, permits of making spaces at will.

206. The Carriage Mechanism. The spring that pulls the typewriter carriage beyond the point of working is wound up by pushing the carriage to the right by the hand. The handle for doing this is connected to a set of levers acting upon toothed wheels which move the circular rubber roller or platten, on which the paper is carried, around sufficiently far to bring the writing being done below that previously finished. By a system of ratchets working on toothed wheels, each stroke of any type-bar moves through a small distance a train of gear-wheels that carry an inked ribbon across the space into which the type-bars strike. For convenience in working, there are not as many type-bars and keys as characters. One type-bar carries two characters and in a few makes of machines, three. Levers attached to the shift keys move either the fulcrums of the type-bars on the carriage so that one or the other of the characters of any type-bar will strike the paper.

By means of various combinations of levers such devices as a bell to indicate when the end of a line of writing is reached, the adjustment of the position of the paper, the using of one edge or another of the ribbon inked with two colors of ink, the going back one or more spaces to correct an error, and so forth, are worked. Perhaps one of the most ingenious of these devices is the combination of four to seven levers of various classes that make the ribbon move over the point of work when a key is struck, and pulls it back between each stroke, so that the work may be seen or corrections made by hand.

In considering the typewriter it should be borne in mind that the amount of work that has to be done by any given lever system is relatively small and consequently in most of its parts levers of the third class are used, giving speed and distance at the expense of power.

207. Special Typewriter Devices. There are many variations on this scheme of connected levers. A recent one is the provision of electric power to work the type-bars. In this case the key and its lever simply push the necessary bit of metal into place to cause the power to be applied to the particular type-bar wanted.

Another ingenious device is the use of an additional lever on each type-bar, with a weight at its end. As the type strikes the paper, the inertia of the weight causes the lever to move so that the force of the impact is lessened and the noise of operating the

machine is greatly reduced. This constitutes a so-called noiseless typewriter. Another variety of typewriter, instead of using type on type-bars, has all the type on a curved carrier. The key levers move this carrier to the place at which the desired letter is to be printed, and then actuate a small hammer carried within the platten. This hammer pushes the paper against the type, which is in position. This arrangement permits of changing the style and size of type used on a machine as often as desired. A machine of this sort is useful, for example, when it is desired to change from Latin to Greek or German script.

208. The Automobile. One of the most complicated and at the same time one of the most effective and commonly used machines today is the automobile. In this machine we have practically every form and class of lever, the wheel and axle, the screw, the inclined plane (exemplified in the blades of the fan), as well as the gas engine with its various accessories, and the electric motor. To understand the working of the automobile even in an elementary sense one must not only know something of the principles of mechanics and thermo-dynamics, but one must possess a fair knowledge of electricity as well. From an educational viewpoint the most significant thing about the automobile is not its wonderful development in recent years, the great perfection of its parts, or its amazing distribution and use; the significant thing is that the introduction of this popular means of locomotion has demonstrated the fact that thousands of people, men and women, and children too, for that matter, possessed a mechanical skill and understanding which hitherto had been entirely unsuspected. Many a boy who in school has struggled with indifferent success over the simplest problems of mechanics has later on surprised his teachers, and himself, by his ability to master the complicated mechanism of a newly acquired automobile.

The power developed by the automobile engine is delivered to the rear wheels through what is known as the power transmission system, Fig. 189, which consists of the clutch, transmission gears, universal joints, drive shaft, differential, and rear axles. We shall here consider the application of the simple mechanical powers to the automobile, leaving for later chapters a discussion of the gas engine and the necessary electrical appliances.

209. Clutch. In order to develop power an automobile engine must attain a relatively high speed. It is therefore necessary to disconnect the engine from the power transmission system while

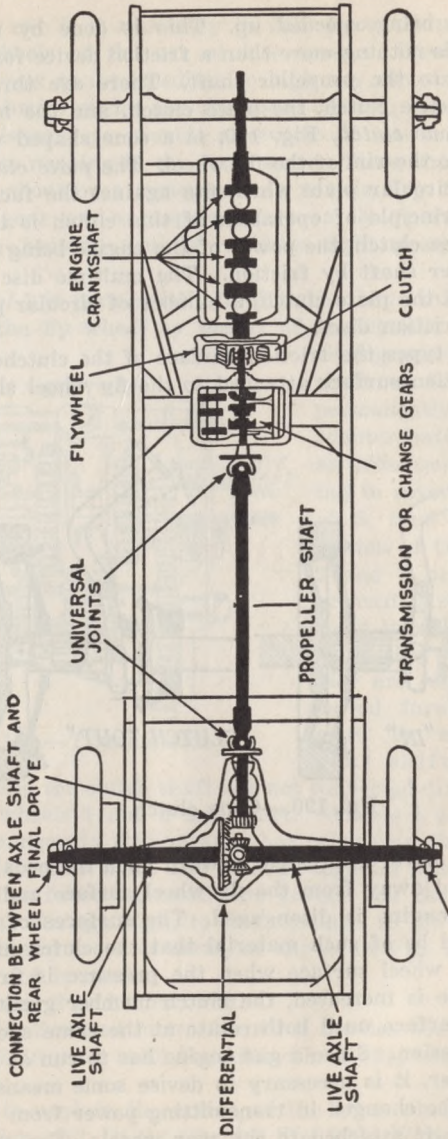


FIG. 189.—Automobile power transmission system

the engine is being speeded up. This is done by means of a clutch, which is nothing more than a friction device for connecting the fly wheel to the propeller shaft. There are three types of clutches—the cone clutch, the plate clutch, and the multiple disc clutch. The *cone clutch,* Fig. 190, is a cone shaped wheel which fits snugly into the rim of the fly wheel. The *plate clutch* consists of a friction circular plate which fits against the face of the fly wheel. The principle of operation of this clutch is the same as that of the cone clutch, the power of the engine being transmitted to the propeller shaft by friction. The multiple disc clutch is a modification of the plate clutch, a number of circular plates rotating between friction discs.

In all three types the friction surface of the clutches is forced against a friction surface attached to the fly wheel shaft. When

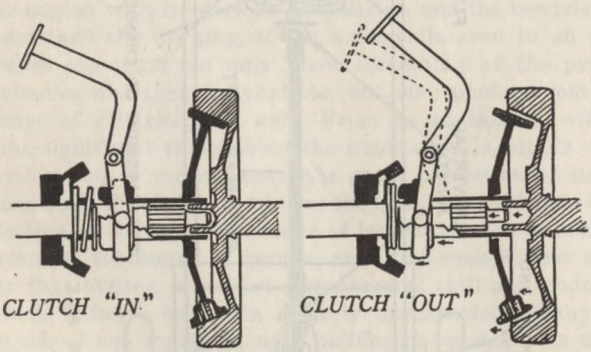


FIG. 190.—Cone clutch

the foot of the driver is pressed forward upon the pedal lever, the clutch is pushed away from the fly wheel surface and the power system of the engine is disengaged. The surfaces of the clutch members should be of such material that the clutch member will slip on the fly wheel surface when the pressure is first applied. As the pressure is increased, the clutch member gradually grips the fly wheel surface until both rotate at the same speed.

210. Transmission. Since a gas engine has to run at high speed to develop power, it is necessary to devise some means by which the speed may be changed in transmitting power from the engine to the drive shaft attached to the rear wheels. For example, in

starting the car it is necessary to deliver considerable power to the drive wheels of the car at low speed. Likewise successive speeds from low to intermediate to high are necessary depending upon the highway grades and the speed desired. Also, to back the car it is necessary to provide a reverse of the drive shaft, the engine always running in the same sense. These changes in speed are brought about by a system of cog wheels, Fig. 191, known as the change gear set, or appropriately as the *transmission*. The transmission housing usually contains eight geared wheels, as follows:

Gear wheel 1 is permanently fixed to the drive shaft which connects with the fly wheel by means of the clutch; 2 and 3 are sliding gears on the drive shaft; 4, 5, 6, and 7 are permanently fixed to the countershaft; 8 is called an idler gear, its use being to reverse the motion of 3, thus reversing the motion of the car. Gears 2 and 3 are set upon a grooved part of the drive shaft in such a way that while they rotate with it they are capable of being moved forward or backward by means of the gear shift lever. The drive shaft and the clutch shaft are not connected directly together when gear wheels 1 and 2 are apart. Wheels 1 and 4 are permanently connected; this means that whenever the drive shaft rotates, the counter shaft also rotates, together with wheels 5, 6, 7, and 8. The following conditions govern the changes in speed from neutral through high, to intermediate, to low:

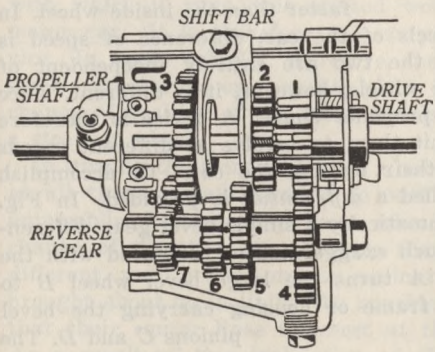


FIG. 191.—Transmission

Neutral, 2 and 3 out of mesh, as shown in Fig. 191.

Low, 3 and 6 in mesh.

Intermediate, 2 and 5 in mesh.

High speed, 2 moved forward until it meshes with 1.

Reverse, 3 in mesh with the idler wheel (8), which is in mesh with 7.

In order to explain the action of the idler wheel 8 in reversing the motion of 3, let us consider Fig. 192. When wheels 3 and

6 are in mesh and 6 rotates in a clock-wise sense, 3 rotates in an opposite sense. Now suppose that 3 is placed in mesh with 8 and 7. Wheel 6 rotates in a clock-wise sense as before. This communicates a counter motion to 8, which in turn causes 3 to rotate in a sense opposite to that when in mesh with 6.

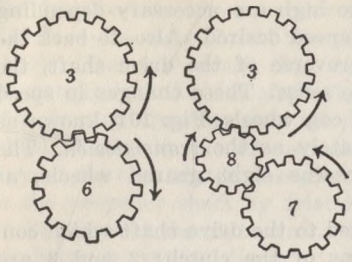


FIG. 192.—Reversing gears

211. Differential. When a car turns a corner, it is evident that the outside wheel must travel faster than the inside wheel. In the case of the front wheels of the car, difference of speed is easily accomplished, since the two are entirely independent of each other. With the rear wheels, however, it is different, since both are connected to the propeller shaft. A device is therefore necessary which will permit them to revolve at different speeds without interfering with their driving the car. To accomplish this a system of gears called a *differential* is provided. In Fig. 193 there is shown diagrammatically a simple bevel gear differential, the size being very much exaggerated as compared with the wheels. The drive pinion *A* turns the large bevel wheel *B* to which is attached a steel frame or housing carrying the bevel pinions *C* and *D*. The gears *E* and *F* are attached directly to the axles marked *R* and *L*. These gears connect indirectly with *B* through the pinions *C* and *D*. With equal resistance on each wheel the whole gear system and the axles rotate at the same speed. When, on the other hand, the car turns a corner or runs over uneven ground, there will be a movement of the differential pinions

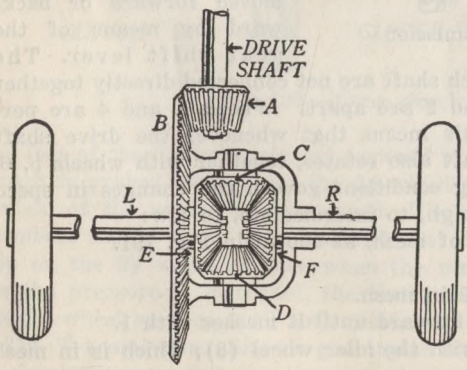


FIG. 193.—Differential

around their axles, thus enabling one car wheel to revolve faster than the other.

There are three general types of differential gears employed on modern motor vehicles. These are the bevel gear, the spur gear, and the worm gear differentials. In general the bevel gear type is used.

212. Steering Gear. In order to change the direction of a moving vehicle, such as a horse drawn carriage or an automobile, it is necessary to change the position of the steering (front) wheels. In a horse drawn vehicle this change is brought about by moving the axle to which the front wheels are attached about a single midway pivot of support. Because of the weight resting on the front axle of an automobile and also because of the speed with which it runs, this method would be impracticable for a motor car. It is therefore necessary to employ some other device. In motor vehicles, therefore, the axle is held rigid and the wheels are turned on separate pivot joints called knuckles. This change of position of the front wheels of the car is controlled by a steering wheel.

In turning the front wheels of a horse-drawn vehicle, the wheels at all times remain parallel to each other. In the case of an automobile, however, the front wheels are not parallel when a change of direction occurs, since they are perpendicular to two different radii. The proper inclination of the two wheels is brought about by inclining the knuckle arms toward each other so that their center lines intersect at the center of the rear axle. As a result of this inclination of the knuckle arms the inside wheels is turned more than the outside one in changing direction.

The change of direction of the car wheels is controlled by a steering wheel, Fig. 194, which is connected at the lower end to the drag rod by means of a worm gear. The two knuckle arms are connected by a tie rod which transmits the turning motion to both wheels. Since the resistance of the road tends to spread the front wheels apart, these wheels are "gathered" slightly; that is, the front edges are slightly nearer together than the rear edges of the wheels.

213. Brakes. The brakes constitute a very important part of an automobile. A brake consists of two essential parts, a brake drum and a brake band. In some cases the brake band passes around the outside of the drum and in other cases inside. In both types, however, the band exerts friction on the drum when the brakes are set. Since an unequal grip of the brake bands

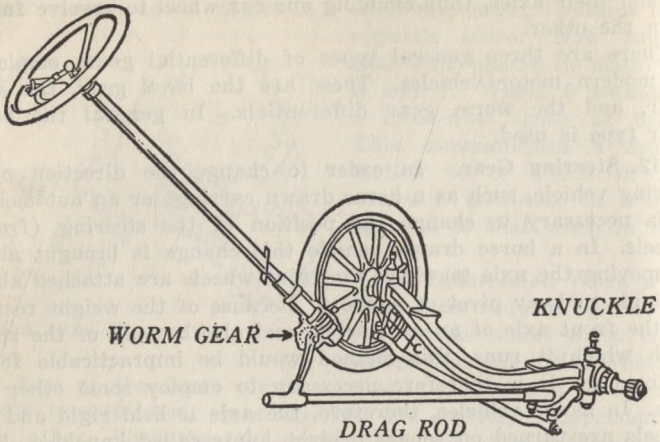


FIG. 194.—Steering gear

upon the wheels would tend to make the car skid, some method of equalizing the pull transmitted from the brake pedal or lever to the drum is necessary. This is accomplished by means of brake equalizers. Most cars are now equipped with four-wheel brakes.

SUMMARY

1. Work = force × space = Fs.

2. Gravitational units of work = foot pound, gram centimeter, kilogram-meter.

(a) Foot pound = 1 pound × 1 foot
(b) Gram centimeter = 1 gram × 1 centimeter
(c) Kilogram-meter = 1 kilogram × 1 meter

3. Absolute units of work = foot poundal, erg, joule.

(a) Foot poundal = 1 poundal × 1 foot
(b) Erg = 1 dyne × 1 centimeter
(c) Joule = 10^7 ergs.

4. Power = work per unit of time = $\frac{work}{time} = \frac{w}{t}$.

5. Horse power = 33,000 foot pounds per minute

$$= \frac{foot\ pounds}{33{,}000 \times minutes} = \frac{foot\ pounds}{550 \times seconds}$$

6. Watt = 10^7 ergs per second = 1 joule per second.

7. 1 horse power = 746 watts.

8. Energy is capacity for doing work. Units of energy (potential and kinetic) same as for work.

9. Potential energy $= Fs$.

10. Kinetic energy $= \frac{1}{2}\ mv^2$, *absolute units.*

$$= \frac{1}{2}\ \frac{mv^2}{g}, \textit{ gravitational units.}$$

11. Moment of a force = force × lever arm $= Fd$.

12. Simple machines (mechanical powers): lever, wheel and axle, pulley, inclined plane, wedge, screw.

13. Law of ideal frictionless machine: Applied force (effort) × space through which it acts = resistance overcome × space through which it acts; that is, $Fs = Rs'$. The factors s and s' may be expressed in terms of certain working parts of the machine.

14. Mechanical advantage $= \frac{R}{F} = \frac{s}{s'}$

15. Mechanical efficiency is

$$M.\ E. = \frac{\textit{useful work done by the machine}}{\textit{work done on the machine}} = \frac{\textit{work out}}{\textit{work in}}.$$

Work in = work out + frictional losses.

16. Friction = resistance encountered when one body moves over another. Friction may be sliding (kinetic) or rolling.

17. Coefficient of kinetic friction is the force required to keep a body in uniform motion divided by the normal force pressing the bodies together; that is,

$$\textit{coefficient kinetic friction} = \frac{\textit{frictional force}}{\textit{normal force}} = \frac{F}{N}$$

REVIEW EXERCISES

A

1. In the construction of a building a hoisting engine lifts 23,100 lb. of brick to a height of 20 ft. in 2 minutes. On another building a man carries the same quantity of brick to a height of 20 ft. in 1 hour and 40 minutes. Considering the brick lifted in each case, how much work in foot pounds was done by (a) the engine? (b) the workman?

2. At what rate in h. p. did (a) the engine work? (b) the man?

3. A man weighing 180 lb. and a boy weighing 98 lb ascend

the stairs of an office building to a height of 55 ft. The man walks up in 3 minutes and the boy runs up in 30 seconds. At what rate in h. p. did each work in ascending the stairs?

4. A stone having a mass of 2 lb. is thrown to a height of 20 ft. At what point does it have (a) its greatest kinetic energy? (b) its greatest potential energy? (c) What is its potential energy in foot pounds at the upper limit of its flight?

5. (a) How does the P.E. which the stone (Ex. 4) possesses at its highest point compare with the K.E. it will have when it returns to the place from which it was thrown? (b) What is this K.E. in foot pounds? foot poundals?

6. How may the stability of a body be increased? Explain by means of a sketch why a rectangular body lying on its side is more stable than when standing on end.

7. By means of a sketch explain what you mean by *moment of a force.* Assign numerical values to F and d, and compute the moment about an axis O.

8. The dimensions of two doors, A and B, both made of the same material, are 6 x 10 ft. Door A is hinged on the side, and B on the end. In which case should the hinges be the stronger, and why?

9. Two men, A and B, carry a load weighing 240 lb. suspended from a bar 6 ft. in length. The weight is 2 ft. from A. Neglecting the weight of the bar, find how much of the load each man carries, assuming that the men are located at the ends of the bar.

10. Make sketches to illustrate a lever of (a) the first class, (b) second class, (c) third class. State clearly the position of the fulcrum in each case with reference to F and R.

11. Suppose you have a rigid bar of given length for use as a lever. In which case will it give the greater mechanical advantage, when used as a lever of the first class, or second class? Explain fully.

12. What advantage is obtained by means of third class lever? Give an example.

13. Draw three pulleys as shown in Fig. 170 C and find what force will be required to support a weight of 600 lb. What is the mechanical advantage in this case?

14. Draw four pulleys as shown in Fig. 170 D and find what force will be required to support a weight of 600 lb. What is the mechanical advantage?

15. It is desired to lift vertically to a height of 4 ft. a cake of ice weighing 420 lb., by sliding it up a 12 ft. plank used as an inclined plane. (a) Neglecting friction, what force is required to slide the ice up the plane? (b) How much work is done? (c) What is the mechanical advantage?

16. A man lifts a weight by means of a jack screw, the distance between the threads of which is ¼ in.; that is, for every turn of the screw the weight is lifted ¼ of an inch. The lever arm of the screw is 3 ft. in length. What resistance is the man capable of overcoming if he exerts a force of 100 lb.? What is the mechanical advantage?

17. (a) On what does the fact that a trolley-car can move along the track without cogs in the track or other projections to push against depend? (b) Railway locomotives are provided with a sand box from which tubes lead down to and in front of the drive wheels. What is the object of sanding the rails, and when is it done?

B

18. Two automobiles of equal weight, 2500 lb., run at uniform rates to the top of two hills, *A* and *B*. The slope distance from the bottom of hill *A* to the top is half a mile, and that of hill *B*, one mile. The vertical height of each hill is 220 ft. Neglecting friction, how much work was done against gravity by each automobile?

19. A farmer is plowing with a tractor which is traveling at the rate of 2 mi. per h. If the average pull on the draw bar of the plow is 1800 lb., at what average rate (h. p.) is the tractor working?

20. A truck horse is pulling a load weighing one ton at the rate of 1.5 mi. per h. The pull exerted by the horse is 180 lb. At what rate in h. p. is the horse working?

21. The power of a street car motor was rated at 90 kilowatts. What was its horse power?

22. A weight of 15,600 kilograms is lifted to a height of 10 meters in 2 minutes. Find the rate at which the work is done in (a) watts; (b) kilowatts; (c) horse power.

23. An iron safe weighing 1.25 tons is pulled up heavy skids onto a truck 3 ft. in height by a force of 600 lb. Find the length of the skids, provided friction is not taken into account.

24. A jack screw, having two threads to the inch, is used to

lift a weight of 2.5 tons. The force arm d of the screw is 2 ft. in length. Find the force required, assuming that the frictional forces are equivalent to 200 lb.

25. In raising a block of granite weighing one and a quarter tons to a height of 80 ft. a hoisting engine does 330,000 foot pounds of work (useful work + frictional loss). Find the efficiency of the system.

26. An electric motor, the energy input of which is 40 kilowatts (kw), delivers 45 h. p. What is the efficiency of the motor?

27. A centrifugal pump which is driven by a 120 h. p. motor delivers 40,000 gallons of water per minute to a height of 10 ft. (a) What is the work done by the pump, taking the weight of a gallon of water to be 8 lb.? (b) What is the efficiency of the pump?

28. A steel block slides over a brass plate. The coefficient of friction of steel on brass is 0.19. The weight of the steel block is 12 lb. What force applied parallel to the surface of the brass plate will be required to pull the steel block along with uniform speed?

C

29. Referring to the laws of falling bodies (Art. 130), find the velocity that the stone (Ex. 4) will have when it strikes the ground. Having now the mass of the stone and its velocity, compute its K.E. by means of the equation $K.E. = \frac{1}{2}\ mv^2$. Compare this result with that found in Ex. 5.

30. A ball having a mass of 1 oz. is fired from a gun of mass 10 lb. with a velocity of 1600 ft. per second. What is the K.E. of the ball (a) in foot poundals? (b) foot pounds?

31. (a) Find the velocity of the gun's recoil. Find the K.E. of the gun. (b) The K.E. of the ball is how many times as great as that of the gun?

32. Count the number of cogs on each of the three geared wheels shown in Fig. 183. (a) How many revolutions of the small wheel will be necessary to produce one revolution of the large wheel? (b) What is the mechanical advantage of this machine, assuming that the force F is applied to the small wheel?

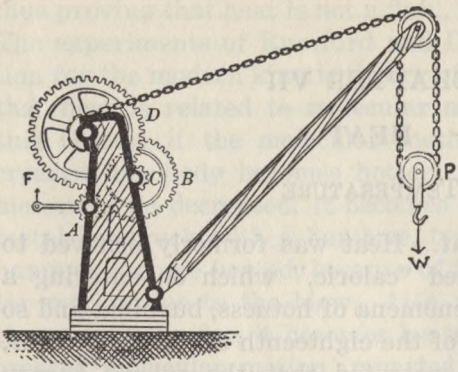


FIG. 195

33. The number of cogs on the gear wheels of the hoisting crane shown in Fig. 195 is as follows: Wheel *A*, 12 cogs; *B*, 48; *C*, 12; and *D*, 60. The radius of the axle of wheel *D* is 2 inches. Wheel *A* meshes into *B*, and *C* into *D*. The length of the crank arm *F* is 18 inches, measured from the center of wheel *A*. (a) Find the mechanical advantage of this machine, taking into account the movable pulley *P*. (b) Neglecting friction, what weight *W* can be lifted by a force of 120 lb.?

34. A small toothed wheel having 20 cogs is attached directly to the armature shaft of a street car motor. This wheel (20 cogs) is geared to a larger cog wheel which has 36 cogs and which is attached to the driving axle of one of the car wheels. The diameter of the car wheel is 32 inches. How many revolutions will the motor have to make in order to drive the car forward a distance of one mile?

35. A workman in grinding a tool presses the tool against the rotating surface of the grindstone with a force of 20 lb. The lateral pull on the tool due to the rotation of the grindstone is 14.4 lb. What is the coefficient of friction between the tool and the stone?

NOTE: For additional problems, see Appendix.

CHAPTER VII

HEAT

TEMPERATURE

214. Nature of Heat. Heat was formerly believed to be a subtle fluid called "caloric," which on entering a body produced the phenomena of hotness, burning, and so forth. About the end of the eighteenth century, however, Count Rumford demonstrated that the caloric theory (fluid theory) of heat is untenable. During the supervision of the boring of some brass cannon, he noted that there was some relation between the work done by the horse driving the machinery and the heat developed by the boring tool. At this time it was supposed that heat was a fluid. Rumford decided that there must be a definite relation between the motion of the particles of brass and the heat generated. He went so far as to measure the quantity of heat given off in the boring of a cannon, and in this experiment he came very near to discovering the mechanical equivalent of heat.

Count Rumford (Benjamin Thompson) was one of the most interesting characters of his day. He was born in Massachusetts in 1753. He went to England during the Revolutionary period, and later went to Germany where he settled at Munich, in the employ of the elector of Bavaria, with whom he acquired great influence. On the occasion of receiving the title of Count, in 1790, he chose the name Rumford after the village in New Hampshire where he had resided prior to leaving for England.

In 1799 Sir Humphrey Davy gave the death blow to the caloric theory by showing that two pieces of ice can be

liquefied at the freezing point by rubbing them together, thus proving that heat is not a fluid, but a form of energy. The experiments of Rumford and Davy laid the foundation for the modern kinetic theory of heat, which assumes that heat is related to molecular motion. According to this theory, if the molecular motion of a body be increased, the body becomes hotter; if the motion of the molecules be decreased, it becomes cooler. If a piece of metal be struck with a hammer, both the metal and the hammer become heated, because of the increased molecular motion due to the blow. Also, if a piece of iron be thrust into the fire, it becomes heated, because of the increased molecular motion imparted by the flame.

Heat may be defined as a form of energy due to the molecular motions of a body.

The principal sources of heat are: (a) the heavenly bodies, such as the sun and the stars; (b) the interior of the earth; (c) chemical action, as in the combustion of wood or coal; (d) electrical energy, as in the heating of the filament of an incandescent lamp; and (e) mechanical sources, such as friction, impact, and compression.

215. Temperature. The *temperature* of a body is that which determines its degree of hotness or coldness. We speak of boiling water as having a high temperature, and of ice-cold water as having a low temperature. The temperature of a body is determined by the average kinetic energy of its molecules; the greater the molecular motion the higher the temperature.

We depend in a great many cases upon our sensations of hot and cold to determine the relative temperatures of bodies. In this manner we may determine that a steam pipe is hot and that ice is cold. This primitive method of determining temperature, however, is not reliable, as may be illustrated by a simple experiment. Take three basins of water, one hot, one cold, and one tepid. Place

one hand in the hot water and the other in the cold and hold them there for a few moments. Now place both hands in the tepid water. The hand that was in the hot water will feel cool and the hand that was in the cold water will feel warm. Thus the sensations of hot and cold may give us unreliable information as to the actual temperature of the water in the third basin. It is therefore necessary to have some more satisfactory means of measuring temperature, and for this purpose we use the thermometer.

216. The Mercury Thermometer. A *thermometer* is an instrument for measuring temperature. The measurement of temperature by means of the mercury thermometer depends upon the principle that "heat expands, and cold contracts." The expansion of a substance due to heat may be illustrated by means of a piece of apparatus as shown in Fig. 196. A small flask having a glass tube thrust into the rubber stopper is filled with colored water. If now the bulb be put into a beaker of hot water two things may be observed: (a) First, the liquid in the tube falls slightly and then (b) it rises rapidly to a given height. The liquid falls at first because of the initial expansion of the glass bulb; the heat, however, soon affects the liquid, which, expanding more rapidly than the glass, rises in the tube.



FIG. 196
A water thermometer

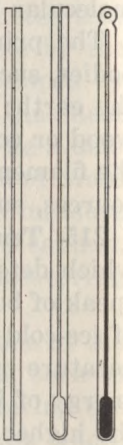


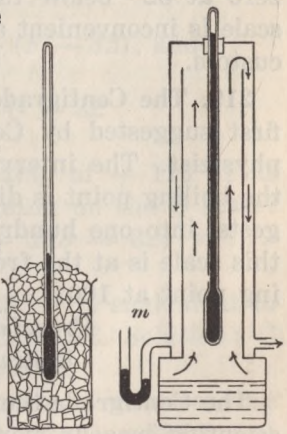
FIG. 197
Showng how a thermometer is made

The mercury thermometer consists of a glass tube containing mercury as the expanding substance. The

thermometer is constructed and filled as follows: On the end of a glass tube having a capillary bore *a*, Fig. 197, there is blown a thin-walled bulb *b*. This bulb and a short portion of the capillary are filled with mercury by expelling the air from the bulb by heating, and then thrusting the open end of the tube into the mercury, which rises in the tube as the air in the bulb cools and contracts. The instrument is now inverted and heated until the mercury expands sufficiently to fill the entire bore, at which time the tube is sealed off at the top, as shown in *c*, thus leaving it air tight. When the thermometer cools the mercury contracts into the lower part, leaving the major portion of the capillary free.

217. The Fixed Points on a Thermometer. In order to establish a thermometer scale, it is necessary to determine two fixed points. These points are the freezing and boiling points of water under a pressure of one atmosphere. The *freezing point* on the scale is determined by placing the bulb and part of the stem in finely crushed ice, Fig. 198 *I*. The mercury in the bulb contracts, and hence the thread of mercury in the capillary falls. The point at which it comes to rest is marked. This is called the freezing point; it is the temperature at which water freezes or ice melts, the two points being practically the same.

The *boiling point* is determined by suspending the thermometer in steam arising from water boiling under a pressure of one atmosphere. The apparatus used in de-



I *II*

FIG. 198.—Freezing and boiling points

termining the boiling point is shown in Fig. 198 *II*. The required pressure of one atmosphere is determined by means of the manometer gauge *m*. As the thermometer becomes heated by the steam which is of the same temperature as the boiling water, the mercury rises to a certain point, where it remains stationary. A mark is made on the scale to indicate this point, which is called the boiling point.

218. The Fahrenheit Scale. This scale was first used by Fahrenheit, a German scientist. The interval between the freezing point and the boiling point is divided into 180 degrees, usually written 180°. The zero of the scale is set at 32° below the freezing point, thus making the interval from the zero of the scale to the boiling point 212°.

$$\textit{Freezing point} = 32°\ F.$$
$$\textit{Boiling point} = 212°\ F.$$

There is no advantage, as we now know, in having the zero at 32° below the freezing point. The Fahrenheit scale is inconvenient and remains in use largely through custom.

219. The Centigrade Scale. The Centigrade scale was first suggested by Celsius, a Swedish astronomer and physicist. The interval between the freezing point and the boiling point is divided, as the word centigrade suggests, into one hundred grades or degrees. The zero of this scale is at the freezing point of water, and the boiling point at 100°.

$$\textit{Freezing point} = 0°\ C.$$
$$\textit{Boiling point} = 100°\ C.$$

The Centigrade thermometer is used in practically all scientific measurements of temperature and is, on account of the simplicity and convenience of the scale, coming more and more into common use.

220. Comparison of Centigrade and Fahrenheit Scales. In Fig. 199 the two scales are drawn side by side for comparison. Since the interval between the freezing point and the boiling point is the same for both scales, differing only in the number of units marked on each, and since 100 C units are equal in length to 180 F units, it follows that for a given length of scale

$$\frac{C \text{ units}}{F \text{ units}} = \frac{100}{180} = \frac{5}{9}$$

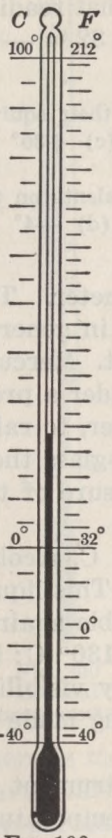


FIG. 199 Centigrade and Fahrenheit scales

Now, bearing in mind that to change from one scale to the other it is necessary to reckon from some fixed point such as the freezing point and to do this we must always take into account the 32 units between zero and the freezing point of the Fahrenheit scale, to change temperature readings from one scale to the other we may use the equations

$$C = \frac{5}{9}(F - 32), \text{ and}$$

$$F = \frac{9}{5}C + 32$$

Examples. 1. A reading of + 113° F is equivalent to what reading on the C scale? *Solution*: $C = 5/9\ (+ 113 - 32) = 5/9\ (+ 81) = + 45°$C.

2. A reading of + 23° F is equivalent to what reading on the C scale? *Solution*: $C = 5/9\ (+23 = 32) = -5°$C.

3. A reading of — 13° F is equivalent to what reading on the C scale? *Solution*: $C = 5/9\ (- 13 - 32) = 5/9\ (- 45) = -25°$ C.

4. A reading of +20° C is equivalent to what reading

on the F scale? *Solution*: $F = 9/5\ (+20) + 32 = +36 + 32 = +68°$ F.

5. A reading of $-20°$ C is equivalent to what reading on the F scale? *Solution*: $F = 9/5\ (-20) + 32 = -36 + 32 = -4°$ F.

EXERCISES. **1.** Give for the following C readings their equivalents on the F scale: (a) $+10°$ C; (b) $-10°$ C; (c) $+30°$ C; (d) $-30°$ C; (e) $-40°$ C.

2. Give for the following F readings their equivalents on the C scale: (a) $+32°$ F; (b) $+77°$ F; (c) $+5°$ F; (d) $-4°$ F; (e) $-40°$ F.

221. Limitations of the Mercury Thermometer. The limitations of a thermometric substance are, in general, fixed by its freezing point and its boiling point. Mercury freezes at $-39°$ C and boils at $+357°$ C, under a pressure of one atmosphere. It is possible, however, to raise the boiling point of mercury in a mercury-in-glass thermometer to $+500°$ C by increasing the pressure of the gas enclosed within the instrument.

For measuring temperatures below $-39°$ C, alcohol is often used as the thermometric substance. This liquid is usually colored red or blue to render it visible against the glass. The freezing point of alcohol is $-130°$ C; its boiling point is $+78°$ C. Because of the easy visibility of the colored liquid, an increasing use is being made of alcohol thermometers.

222. The Clinical Thermometer. This instrument is used by physicians and nurses in taking the temperature of the human body. It is a mercury-in-glass thermometer, Fig. 200, and is characterized by having a narrow constriction *c* in the capillary. When the thermometer cools the mercury thread breaks at this point, leaving the column of mercury in the tube as it was at the highest temperature while in contact with the body. Thus the reading may be made some time after the temperature

is taken. Before this thermometer can be used again, the mercury in the stem must be shaken down into the bulb by giving the instrument a jerking motion.

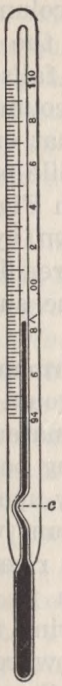


FIG. 200
Clinical thermometer

The normal temperature of the human body is 98.6° F.

223. The Maximum and Minimum Thermometer. It is frequently important to know the highest or the lowest temperature reached during a given period, as for example during the night. For this purpose a *maximum and minimum thermometer* is used, one form of which is shown in Fig. 201. The large bulb *B* and the tube as far as the point *A* contain alcohol or some other suitable liquid. The tube from *A* to *C* contains mercury. Above *C* there is more alcohol, which extends up into *D*, but only partly fills it, thus giving room for expansion. Maximum temperatures are registered on the scale on the right hand tube and minimum temperatures on the left hand. Above the mercury in each tube there are small iron indices, *i* and *i*, which are provided with light friction springs. When the mercury rises in either tube, it pushes up the index, which remains in place when the mercury falls. The positions of the indices indicate the highest and lowest temperatures reached.

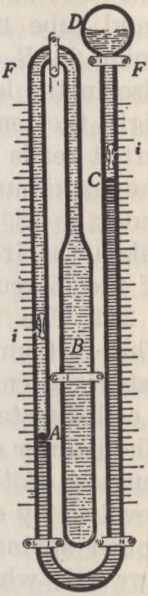


FIG. 201
Maximum and minimum thermometer

When the temperature rises, the liquid in the elongated bulb *B* expands, causing the mercury column to fall at *A* and rise at *C*, pushing the index in the right hand tube upward. When the temperature falls, the liquid in *B* contracts allowing the mercury column to rise in the left hand tube. It will be noted that on the right the temperature scale reads upward, while on the left it reads downward. On the scale shown in Fig. 201 the maximum temperature recorded, as shown by the index, is 72°, and the minimum 44°. After a reading is taken the iron indices are pulled down to the surface of the mercury by means of a small magnet.

224. The Thermograph or Recording Thermometer. This instrument, as the name thermograph would indicate, is a temperature writer. The expanding material is a coiled metal spring attached to which is a long pointer. The pointer carries a pen which presses lightly upon the surface of the record sheet on a rotating drum which revolves by clockwork. When the temperature rises, the coiled spring opens slightly causing the pen to move upward; when the temperature falls, the spring closes causing the pen to move downward. This upward and downward motion of the recording pen together with the rotating motion of the drum makes a continuous record of the temperature, usually covering a period of a week.

Thermographs are used in all Weather Bureau stations and in other places where a continuous record of temperature is desired.

Measurement of Heat

225. Distinction between Temperature and Quantity of Heat. It is important to note that there is a distinction between the temperature of a body and the quantity of heat which it contains. For example, a cup of boiling

water has a temperature very much higher than that of the water in a lake under ordinary summer conditions, yet the quantity of heat in the water of the lake is many times greater than that of the water in the cup. The temperature of a body depends on the average kinetic energy of its molecules; the quantity of heat in the body depends not only upon its temperature but also upon its mass.

226. Unit Quantity of Heat. There are two units employed in the measurement of heat: (a) the calorie, based on the gram mass and the Centigrade scale, and (b) the British thermal unit (B. t. u.), based on the pound mass and the Fahrenheit scale.

A *calorie* is the quantity of heat required to raise the temperature of one gram of water one degree Centigrade. The quantity of heat, for example, required to raise one liter of water (1000 grams) from 20° C to the boiling point (100° C) is $1000 \times (100 - 20) = 80{,}000$ calories. The *large calorie* is 1000 calories; that is, it is the amount of heat required to raise the temperature of 1 kilogram of water 1° C. This is the unit usually used in giving the energy which foods supply, and when used for this purpose in diatetic tables it is usually called "calory," without the prefix "greater."

A *British thermal unit* (B. t. u.) is the quantity of heat required to raise the temperature of one pound of water one degree Fahrenheit. The number of B. t. u. required to raise one gallon of water (about 8 pounds) from 70° F to the boiling point (212° F) is $8 \times (212 - 70) = 1136$ B. t u. This unit is used mainly in giving the heating power of fuels.

227. Specific Heat. All bodies do not have the same capacity for heat. If we take one pound of water and one pound of iron and raise the temperature of each one degree, it will be found that while the water will require

one unit of heat the iron will require only about one-tenth of a unit. The *specific heat* of a substance is numerically equal to the number of units of heat required to raise the temperature of unit mass one degree. The specific heat of water is 1; that is, it requires 1 calorie to raise 1 gram 1° C; or in English units it requires 1 B. t. u. to raise 1 pound 1° F. We then have

$$\textit{specific heat} = \frac{\textit{heat units lost or gained}}{\textit{mass} \times \textit{change of temperature}}$$

in which t is the *change* of temperature. We then may write for the heat gained or lost

$$H = \textit{mass} \times \textit{change of temperature} \times \textit{specific heat}$$
$$= m \times t \times s$$

When, for example, we say that mercury has a specific heat of 0.033, we mean that it requires 0.033 of a unit (calories or B. t. u.) to raise the temperature of unit mass (gram or pound) one degree (Centigrade or Fahrenheit). The specific heats of a few common substances are given below. For a more complete list, see Specific Heat Table X, Appendix.

SPECIFIC HEATS

Mercury	0.033	Iron	0.113
Silver	0.056	Ice	0.5
Copper	0.09	Water	1.0

228. **How to Find the Specific Heat of a Body.** One of the simplest means of determining the specific heat of a body is known as the method of mixtures. A hot body is brought into contact with a cool body; the hot body loses heat and the cool body gains heat. Suppose that we wish to find the specific heat of lead by the method of mixtures. Suspend a given mass of lead in boiling water until the metal comes to the temperature of the boiling water. The temperature of the water should be deter-

mined by means of a thermometer. Have at hand a copper beaker (calorimeter), containing a known mass of water. The mass of the calorimeter must also be known. Determine the temperature of the water in the calorimeter and then quickly transfer the lead from the hot water to the calorimeter. Stir gently with a thermometer until the mixture in the beaker (lead and water) comes to a constant temperature, which should be noted. We now have the following data: (a) The mass of the lead and its change (fall) in temperature; (b) the mass of the water in the calorimeter and its change (rise) in temperature; (c) the mass of the calorimeter and its change (rise) in temperature. From these data we may compute the specific heat of the lead.

The calculation of the specific heat of a body by the method of mixtures is based on the assumption that the heat lost by one body is equal to the heat gained by other bodies; that is, in the particular case under consideration, the heat lost by the lead is equal to the heat gained by the water, plus that gained by the calorimeter. Now the heat lost or gained by a body, measured in calories or B. t. u., is equal to the mass of the body times its change of temperature times its specific heat; therefore, we may write

$$\textit{heat lost by lead} = \textit{heat gained by water} + \textit{heat gained by calorimeter}$$

$$m \times t \times s = m' \times t' \times s' + m'' \times t'' \times s''$$

where m is the mass of the body losing heat, t is its change in temperature, and s its specific heat; m' is the mass of the water, t' its change in temperature, s' its specific heat, which is equal to unity; m'' is the mass of the calorimeter, t'' is its change in temperature (the same as that of the water), and s'' its specific heat.

Example. Suppose that an iron ball of mass 52 grams

having a temperature of 100° C is dropped into 200 grams of water at 20° C, contained in a copper calorimeter having a mass of 100 grams. The resulting temperature of the calorimeter and its contents is 22.2° C. The specific heat of copper is 0.09. Find the specific heat of the iron. *Solution*: In making a calculation of this sort we assume that the heat lost by the iron is gained by the water and the calorimeter together; therefore, we may write

$$m \times t \times s = m' \times t' \times s' + m'' \times t'' \times s''$$
$$52 \times 77.8 \times s = (200 \times 2.2 \times 1) + (100 \times 2.2 \times 0.09)$$
$$s = 0.113$$

EXERCISES. **3.** A piece of nickel having a mass of 114 grams at 100° C is dropped into 100 grams of water at 10° C. The resulting temperature of the water is 20° C. Find the specific heat of the nickel, neglecting the heat lost to the containing vessel.

4. One kilogram of water and one kilogram of iron, each at a temperature of 100° C, are cooled to 20° C. How much heat in calories is given out by each?

5. A person drinks 300 grams of ice water which comes to the temperature of the body (98.5° F). How many calories of heat are taken up by the water?

Begin March test

EFFECTS OF HEAT

229. Introductory. (a) *Sensations of warmth and cold.* The most noteworthy effect of heat is that experienced in the effects produced in our bodies in the sensation of warmth, and in the absence of heat, the sensation of cold. Bodily heat is one of the prime necessities of life. The problem of regulating the natural heat of the body by means of proper clothing and of heating our houses in cold weather is therefore of primary importance. A discussion of combustion, fuels, and methods of heating will be considered under a later topic. (b) *Expansion.* In general heat causes solids, liquids, and gases to expand; cold, on the other hand, causes them to contract. There

are some exceptions to this rule as applied to solids and liquids. (Art. 234.) (c) *Fusion.* Heat causes solids to fuse or melt. (d) *Vaporization.* Heat causes solids and liquids to vaporize.

230. Expansion of Solids. The expansion of a metal may readily be demonstrated to a class by means of the ball and ring experiment. The brass ball, Fig. 202, when cold will just pass through the ring. Heat the ball for a few moments in a Bunsen flame. The ball expands and now will not pass through the ring. Telegraph and telephone wires in winter are stretched taut from pole to pole, due to contraction. In summer, when they are exposed to the fierce rays of the sun, they expand and sag. If the wires were stretched taut in the summer, there would not be sufficient leeway for the contraction which accompanies cold weather, and they would snap under the strain.

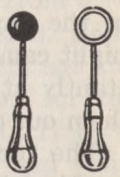


FIG. 202
Ball and ring experiment

The snug fit of metal tires and bands for wheels is due to the contraction resulting from cooling. A metal tire designed to be "shrunk" on, is made slightly smaller than the wheel which it is to protect; the tire is then heated until it has expanded sufficiently to slip on to the wheel. As the tire cools it contracts.

In building a railroad, spaces are usually left between consecutive rails in order to allow for expansion during the summer. The unsightly cracks and humps in cement floors are sometimes due to the expansion resulting from heat. Cracking from this cause can usually be avoided by cutting the soft cement into squares, the spaces between giving them opportunity for expansion just as do the spaces between the rails of a railroad. In the construction of concrete highways, expansion joints are placed every 25 or 30 feet along the road.

Heat plays an important part in the splitting of rocks and in the formation of rock soil. Rocks in exposed places are greatly affected by changes of temperature. In regions where the changes of temperature are sudden and frequent, the rocks are not able to withstand the strain of expansion and contraction, and as a result they crack and crumble. In the Sahara Desert such crumbling of the rock into sand has been caused by the change from the intense heat of the day to the cold of night. The heat of the day causes the rocks to expand and the cold of night causes them to contract, and these two forces constantly at work loosen the grains of the rock and force them out of place, thus bringing about disintegration.

The force exerted by metals when expanding or contracting is usually very great. It is estimated that it would require a pressure of nearly 10,000 pounds per square inch to keep mercury from expanding when heated from 0° C to 10° C. Boiler plates are fastened together by heated rivets; when the rivets cool they contract, forming very tight joints.

231. Coefficient of Linear Expansion. The expansion and contraction due to changes of temperature make it necessary for the engineer to take careful account of the expansion factor in the construction of buildings and bridges, the laying of rails on railroad tracks, and in the installation of water and steam pipes. Suppose, for example, it is desired to put down steel railway rails. One must figure out beforehand just what expansion will take place during the possible extremes of temperature between winter and summer, say from $-20°$ C ($-4°$ F) in winter to $+50°$ C ($+122°$ F) in summer. In order to do this it is necessary to know the rate of expansion of the metal in question. The length of the rate of expansion per degree is called the coefficient of linear expansion, and may be defined thus: The *coefficient of linear expan-*

sion (*C. L. E.*) is the increase in length per degree per unit length, or

$$C.L.E. = \frac{change\ in\ length}{change\ in\ temperature \times original\ length} = \frac{l}{t \times L}$$

where l is the change (increase or decrease) in length, t the change in temperature, and L the original length.

Coefficients of Expansion, Table XI, Appendix.

Example. A metal rod 1 meter in length expands 2 millimeters when its temperature is increased from 0° C to 120° C. Find its coefficient of linear expansion. *Solution*: The original length of the rod was 100 cm and the expansion (increase in length) 0.2 cm. Then for the coefficient of linear expansion we may write: $C.\ L.\ E. = 0.2/(120 \times 100) = 1/60{,}000$ cm per degree per unit length = 0.000017 cm/degree/cm.

EXERCISES. **6.** At a certain place the range of temperature during a given year between winter and summer was —25° C and +40° C. Give the equivalent readings on the F scale.

7. At a temperature of 0° C an iron pipe is 100 ft. long. Its length increases to 100.12 ft. when heated to 100° C by steam passing through it. Find the coefficient of linear expansion of iron.

8. A brass rod having a coefficient of linear expansion of 0.000018 has a length of 180 cm at 0° C. What will be its length at 203° F?

9. A steel rail, such as is used on railroad tracks, 30 ft. long and having a coefficient of linear expansion of 0.000012, will increase in length how many inches when the temperature changes from 0° C to 50° C?

232. Expansion Devices. *The compound bar.* If two strips of different metals, such as brass and iron, be riveted together, Fig. 203, and the compound bar thus formed be heated, it will bend in the form shown in *b*, due to the fact that the coefficient of expansion of brass is greater than that of iron. Some of the applications of the compound bar principle are illustrated in the following topics.

The alarm thermometer. In Fig. 204 there is shown a device for automatically closing an electric circuit by the bending of the

compound bar *B*. If the bar be heated, the brass will expand more rapidly than the iron, and thus cause the bar to bend toward *A*

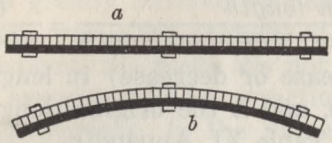


FIG. 203
Compound bar

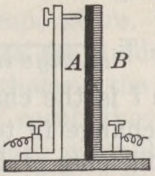


FIG. 204
Alarm thermometer

until electric contact is made. Such a device is sometimes used to give alarm in case of fire.

The compound balance wheel. The balance wheel of a watch serves the same purpose as the pendulum of a clock. If the temperature increases, the length of the spokes change thereby, and also the elasticity of the hair-spring, causing a change in the period of vibration. To offset this change in the rate of vibration the rim of the wheel is made of a compound bar, the metals of which are arranged in such a way that the outer one expands much more rapidly than the inner one, thus causing the ends *A* and *B*, Fig. 205, to bend inward. In this manner the time of vibration of the wheel may be kept practically constant.

Compensating pendulums. We have seen that a change in temperature tends to produce a change in the length of the body. Now, any variation in the length of its pendulum will change the rate of a pendulum clock. Several devices, Fig. 206, have therefore been employed to keep the length constant, one such being the so-called gridiron pendulum, in which the lengthening of the rods marked *s*, usually of steel, tends to lower the bob, while the expansion of the rods *b*, usually of brass, raises the

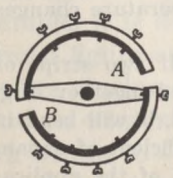


FIG. 205
Compound balance wheel

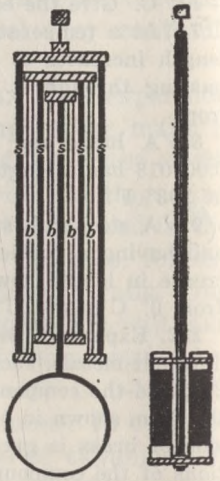


FIG. 206.—Compensating pendulums

bob, and thus the expansion of one set of rods is made to counteract the effect of the other. In the mercury compensating pendulum the lengthening or shortening of the rod is counteracted by a rise or fall of the center of gravity due to the expansion of the mercury in the two glass vessels forming the bob.

The expansion thermometer. The recording pointer of a thermograph (Art. 224) is rigidly attached to the free end of a coiled spring which consists of two metal strips having different coefficients of expansion. The two metals are so chosen that when the temperature rises the increased curvature of the spring causes the pointer to rise, and when it falls the decreased curvature causes the pointer to move downward.

Oven thermometers are of the expansion type. In some cases they are made with a spiral of two metals, as above. In other cases, the expansion of a metal bar is multiplied by levers, the last lever being a pointer which moves across a circular temperature scale.

233. Coefficient of Cubical Expansion. Thus far we have been talking about linear expansion, that is, expansion of a body in one direction. A body, however, may expand in three directions. Consider, for example, a cubical piece of metal 4 centimeters on each edge. It is heated, and as a result, expands in three directions, namely, in length, breadth, and thickness. The *coefficient of cubical expansion* for a given substance is equal to three times its linear coefficient of expansion.

The linear coefficient of expansion of glass, as given in Table XI, Appendix, is 0.000008; its coefficient of cubical expansion is therefore $3 \times 0.000008 = 0.000024$.

EXERCISE. **10.** The volume of a cube of brass at 20° C is 1200 cm^3. Its temperature is raised to 70° C. (a) What is the coefficient of cubical expansion of brass? (b) What is the increase in volume of the brass due to the rise in temperature? (c) What is the volume at 70° C?

234. Expansion of Liquids. Since liquids conform to the shape of the containing vessel, it follows that in determining their coefficients of cubical expansion, account must be taken of the expansion of the vessel as well as

that of the liquid. Like solids, different liquids have different coefficients of expansion. For example, the cubical coefficient of expansion of alcohol is 0.00110; that of mercury, 0.00018.

Water is a partial exception to the rule that liquids expand when heated and contract when cooled. If a quantity of water at 0° C be heated, its volume *decreases* until the temperature of 4° C is reached, after which the volume *increases* as the temperature rises, at 8° C the volume being about the same as at 0° C. At 4° C, therefore, a given mass of water has its least volume, and hence its greatest density. This fact is of great importance in nature, as is illustrated by the freezing of water in a lake in winter. When the temperature falls, the water at the surface reaches its maximum density at 4° C and sinks to the bottom, thus forcing the warm water upward. This goes on until the entire lake reaches a temperature of 4° C, after which the temperature of the water at the surface falls to 0° C and begins to freeze. Since the density of ice is less than that of water at the freezing point, it remains at the surface. The temperature of the water at the bottom of a pond or a lake in winter, is, therefore, 4° C; indeed, the temperature of the deep water of our great lakes remains at about this point throughout the year.

235. Expansion of Gases. With respect to their expansion, gases differ from solids and liquids in two very important particulars. (a) For a given change of temperature, the rate of expansion of gases is much greater than that of solids or liquids under ordinary conditions; (b) all gases have practically the same coefficient of expansion.

The coefficient of cubical expansion of a gas is 1/273. That is, if the temperature of a given volume of gas at 0° C and under constant pressure be increased 1° C, it

will expand 1/273 of its volume; if it be heated to 10° it will expand 10/273, and so on. *Note that the expansion in each case is calculated with reference to the volume at* 0° C. If the gas be cooled, the pressure remaining constant, it will contract 1/273 of its volume at 0° C, per degree. Therefore for gases under *constant pressure* we have

$$\textit{coefficient of cubical expansion of a gas} = \frac{1}{273} = 0.00366$$

236. Absolute Temperature. If a gas enclosed within a vessel of constant volume be heated, the pressure which it exerts will increase; if the gas be cooled, the pressure will decrease. The pressure exerted by a given volume of gas at 0° C is changed 1/273 per degree of change of temperature. Thus it will be noted that the *pressure coefficient* is the same as that of the *volume coefficient*, namely, 1/273 = 0.00366. Now, suppose that the temperature of this gas could be reduced 273° below 0° C; it is evident that its pressure would become zero, which would mean that its molecular motion would also become zero. Since heat is due to molecular motion, it follows that under these conditions there would be no heat. This temperature, 273° below 0° C, is called the *absolute zero*, and a scale based on this zero is called the absolute scale.

Readings on the Centigrade scale may be changed to those on the Absolute scale by simply adding 273 to the C values.

Example. Change the following Centigrade readings to Absolute: (a) 0° C; (b) + 20° C; (c) — 20° C. *Solution*: (a) 0° C = 0 + 273 = 273° *Abs.*; (b) + 20° C = 20 + 273 = 293° *Abs.*; (c) — 20° C = — 20 + 273 = 253° *Abs.*

EXERCISES. **11.** Give the absolute values for the following Centigrade readings: (a) +10° C; (b) —10° C; (c) boiling point on C scale.

12. Give the equivalent Centigrade values for the following readings on the Absolute scale: (a) 243° Abs.; (b) 303° Abs.; (c) 0° Abs.

237. Laws of Charles and Gay-Lussac. The fact that all gases have approximately the same *pressure coefficient* for constant volume was discovered by Jaques Charles, a French physicist; the corresponding fact that all gases have approximately the same *volume coefficient* for constant pressure was discovered a few years later by another French scientist, Joseph Gay-Lussac. These statements of fact are known as the laws of Charles and Gay-Lussac, respectively.

Charles' law. *When the volume is kept constant, the pressure coefficient of all gases is the same, namely,* 1/273 = 0.00366.

Gay-Lussac's law. *When the pressure is kept constant, the volume coefficient of all gases is the same,* 1/273 = 0.00366.

From these laws two very important equations may be derived. (a) From the law of Charles: *When the volume of a gas is kept constant, the pressure is proportional to the absolute temperature; that is*

$$\frac{P}{P'} = \frac{T}{T'}$$

Example. 1. A mass of gas at atmospheric pressure and room temperature (20° C) is enclosed in a vessel so that its volume remains constant. What pressure will it exert on the walls of the containing vessel if the gas be heated to 120° C? *Solution*: A pressure of 1 atmosphere = 1033.3 g/cm². Also, 20° C = 273 + 20 = 293° Abs. and 120° C = 273 + 120 = 393° Abs. Since the atmos-

pheric pressure $P = 1$, the value of P' is therefore

$$\frac{1}{P'} = \frac{293}{393}, \text{ from which}$$

$$P' = 1.3 \text{ } atmosphere's \text{ } pressure = 1343.3 \text{ } g/cm^2$$

(b) From Gay-Lussac's law: *When the pressure remains constant, the volume of a gas is proportional to the absolute temperature; that is,* $V/V' = T/T'$.

Example 2. A mass of gas having a volume of 100 cm^3 at $+20°$ C will have what volume at $-20°$ C, the pressure remaining constant? *Solution*: $+20°$ C $= 293°$ Abs. and $-20°$ C $= 253°$ Abs.; hence

$$\frac{100}{V'} = \frac{293}{253}, \text{ from which}$$

$$V' = 86.3 \text{ } cm^3$$

EXERCISES. **13.** A given volume of gas at 0° C in a steel tank is under a pressure of 40 lb. per sq. in. What pressure will the enclosed gas exert if it be heated to 127° C, assuming that the volume of the tank does not change appreciably.

14. The volume of a given mass of gas under constant pressure and free to expand is 500 cm^3 at 27° C. What will be its volume at —13° C?

15. A gas having a volume of 1000 cm^3 at —10° C is heated and expands under constant pressure to 1200 cm^3. Find the resulting temperature.

238. Standard Hydrogen Thermometer. From the law of Charles we learn that if the volume of a given mass of gas be kept constant, its pressure is proportional to its absolute temperature. This means that temperature may be measured in terms of pressure. The *standard hydrogen thermometer* is constructed to work on this principle. The bulb V, Fig. 207, contains hydrogen gas which exerts pressure upon the mercury at A. This pressure is just counterbalanced by the weight of the mercury column BC. The tube above C is closed and evacuated. By raising or lowering the mercury jar D, the mercury col-

umn at A is kept at a fixed level, indicated by a small pointer inside the tube at A. In this way the volume of the gas in V is kept constant.

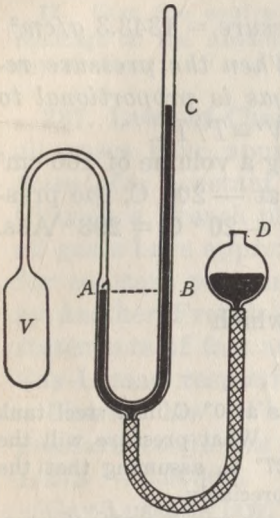


FIG. 207.—Standard hydrogen thermometer

Now if the bulb be placed in ice, the gas in V will contract making it necessary to lower the mercury jar. When the temperature of the gas in V becomes fixed, the pressure registered by the height of the mercury above B will represent that of melting ice, that is, $0°$ C $= 273$ *Abs.* When the bulb V is placed in steam from water boiling under a pressure of one atmosphere, the gas in V will tend to expand. This expansion is prevented by raising the mercury jar. The increased pressure of the gas in V forces the mercury column in BC up until the pressure is again just counterbalanced by the pressure of the mercury column. This pressure corresponds to the temperature of boiling water, $100°$ C $= 373°$ *Abs.* In a similar manner other temperature values may be determined by noting the height of the column in BC.

The hydrogen thermometer is very accurate, but is troublesome to use. Such an instrument is kept at the Bureau of Standards at Washington and is used for calibrating other thermometers when great accuracy is desired.

MELTING, BOILING, VAPORIZATION

239. Fusion and the Melting Point. *Fusion* or melting is the process by which a solid changes to a liquid upon

the application of heat. The melting of ice, sulphur, lead, and tin are all familiar examples of fusion. The temperature at which melting occurs is called the *melting point.*

Many substances like sealing wax, wrought iron, and most kinds of glass have no definite melting points. When heat is applied to such substances, they gradually soften to the point of liquefaction, the change from the solid to the liquid state at no instant being well defined.

Crystalline substances, such as ice, sulphur, cast iron, and most crystalline salts have definite melting points. For example, if a piece of ice at — 10° C be heated, the temperature will rise until it reaches 0° C, at which point the ice will melt. The change from the solid to the liquid state, that is, from ice to water, is abrupt, and takes place at 0° C. Likewise, if water be cooled to this temperature, it freezes. The freezing point of water and the melting point of ice are the same, 0° C, making the dividing line between the solid and the liquid state. Thus, when a piece of ice begins to melt, *the temperature of the ice remains at* 0° *C until it is all melted,* no matter what be the temperature of the water in which it is melting.

240. Heat of Fusion. When a crystalline substance on being heated reaches the melting point, its change from the solid to the liquid state takes place without any rise of temperature. This change, however, requires heat. This heat which is consumed in changing the state of a body without changing its temperature is called the *heat of fusion,* which may be defined as the heat required to change unit mass of a crystalline substance from a solid to a liquid *without changing its temperature.*

The heat of fusion of ice is 80 *calories per gram.* That is, 80 calories of heat are required to change 1 gram of ice at 0° C to water at the same temperature. The heat

of fusion of ice, in English units, is 144 B.t.u. per pound.

241. Change of Volume during Fusion. Substances, in general, contract when they solidify and expand when they melt. Thus if molten lead be poured into a bullet mold and there solidify, it will be found that the bullet does not quite fill the mold; the lead contracts on solidifying and expands on melting.

Certain highly crystalline substances, on the other hand, such as ice, cast iron, and type metal, expand when they solidify and contract when they melt. Such substances as cast iron and type metal are suitable, therefore, for molding, since they expand on solidifying and thus fill every part of the mold, thus reproducing every detail of the pattern.

242. Water Expands on Freezing. It is estimated that 917 cm^3 of water on freezing will become 1000 cm^3 of ice. A cubic centimeter of ice is therefore lighter than a cubic centimeter of water, and hence floats. The pressure exerted by water in freezing is very great, as is often illustrated by the freezing and bursting of water pipes and automobile radiators in winter. A cast iron bomb when filled with water and sealed and placed in a freezing mixture will explode with a loud report. This expansion of water in freezing plays an important part in the disintegration of rocks in the formation of soil.

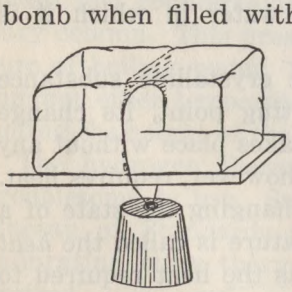


FIG. 208
Melting ice by pressure

243. Relation of Pressure to Fusion. Pressure applied to a substance which *contracts* on melting, as ice for example, lowers the melting point. *Ice therefore may be melted by applying a sufficiently high pressure to it.* Let a fine wire carrying a weight be passed over a small block of ice, as

shown in Fig. 208. The ice just below the wire is melted by the pressure, the water thus formed passing above the wire and freezing again on being released. Thus in a short time the wire will cut its way through the ice, leaving the block as solid as before. ,

If the temperature be near the freezing point, wet snow may be packed into "ice balls," as nearly every schoolboy knows. Also two pieces of ice may be frozen together under warm water by applying considerable pressure and then releasing them suddenly.

The melting point range for different substances is very great, as illustrated by the following comparative values. For a list of melting point values, see Table XII, Appendix.

MELTING POINTS

Ice	0° C	Platinum	1778° C
Lead	327	Iridium	2200
Iron	1200	Tungsten	2950

244. Boiling. If heat be applied to a kettle of water, the following facts may be observed: (a) First, small bubbles of steam form at the bottom, and as they rise, condense, the walls of the bubbles meeting with a sharp impact and thus giving rise to the phenomenon of "singing." (b) When the temperature of the water rises sufficiently high so that the vapor throughout the liquid has a pressure equal to or greater than that of the atmospheric pressure upon the surface, bubbles form copiously at the heating surface and rise to the top of the liquid, producing the agitation called boiling. (c) When the temperature of the liquid reaches the boiling point, it (the temperature) remains constant until the water boils away. Adding more heat causes the water to boil more violently but does not increase the temperature.

Each liquid has a definite boiling point which is constant for constant pressure, Table XIII, Appendix.

Salt dissolved in a liquid raises its boiling point. Most housewives are familiar with the fact that water containing salt boils at a higher temperature than that which is pure. Pure water at a pressure of one atmosphere boils at 100° C. If, however, common table salt (NaCl) be added to the water to the point of saturation, the boiling point may be raised to 109° C.

245. Effect of Pressure on the Boiling Point. The boiling point of a liquid depends upon the pressure exerted upon it. The bubbles that give rise to boiling cannot form in a liquid unless they exert a pressure greater than that of the atmosphere. Therefore, if the pressure upon a liquid be increased, the boiling point is raised. Increasing the pressure raises the boiling point; decreasing the pressure lowers the boiling point.

Put some water at room temperature and atmospheric pressure into a flask, and then connect the flask to an air pump, as shown in Fig. 209. A few strokes of the pump will be sufficient to reduce the pressure so that the water will boil vigorously. This experiment may be performed in another manner, as follows: Heat the water in a flask until its temperature is about 70° C, then cork and invert the flask. Now if water be poured upon it, thus condensing the vapor in the flask and thereby reducing the pressure, the water will boil.

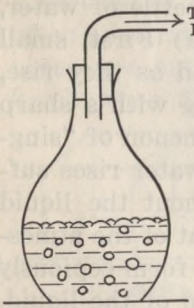


Fig. 209.—Water boiling under reduced pressure

In the manufacture of sugar, vacuum pans are used in which the boiling point of the syrup is lowered by reducing the pressure, thus avoiding the danger of burning the sugar. On the other hand, in the extraction of glue

from bones, the pressure is increased and the boiling point of the liquid is correspondingly raised.

246. Relation of Altitude to Boiling Point. Since atmospheric pressure decreases with the elevation, the boiling point of a liquid also decreases. The boiling point of water may therefore be used to indicate approximately the height of a place above the sea level. A decrease of 1° C in the boiling point indicates an elevation of 295 meters, or about 968 feet.

EXERCISE. **16.** The boiling point of water at sea level is 100° C; on Pike's Peak it is 85.4° C. Compute the approximate elevation of Pike's Peak above sea level from these data.

247. Vaporization. The process by which a substance changes to a vapor is called *vaporization.* It may take place by boiling, by evaporation, or by sublimation. Vaporization takes place at all temperatures, even below the freezing point. It has been found that a block of ice, if left for a few days at a temperature below zero, will lose considerably in weight by vaporization; clothing hung out to dry on a cold winter's day will dry though frozen.

Evaporation is the process by which a liquid changes quietly to a vapor state at the surface.

Sublimation is the process by which a solid changes directly to a vapor without passing through the liquid state. If a piece of ice, for example, be placed under the receiver of an air pump, and the pressure be reduced to 4 millimeters or less, the ice on being heated will not liquefy, but will pass directly from the solid state to a vapor; that is, it will sublime. If the pressure be above 4 millimeters, it will liquefy. Camphor and arsenic will sublime at atmospheric pressure. These substances may be liquefied by sufficiently increasing the pressure upon them. Since arsenic sublimes at atmospheric pressure, it is considered detrimental to health to use wall paper

which contains any considerable amount of arsenic in the coloring matter.

248. Rate of Evaporation. The rate at which evaporation goes on depends upon four factors: (a) The extent of the free surface of the liquid, (b) the temperature, (c) the pressure exerted upon the surface of the liquid, and (d) the degree of saturation of the space above the liquid.

It is a familiar experience that if a given quantity of water be placed in a broad shallow dish, it will evaporate much more quickly than if placed in a deep, narrow dish; and also that the higher the temperature, the more rapid is the rate of evaporation. The drying of roads and streets after a rain illustrates the effect of a change of air on the rate of evaporation. If after a rain there be no wind, the air soon becomes saturated and evaporation is retarded; if, however, the air be in motion, evaporation goes on rapidly and the roads soon become dry. Damp clothes hung out to dry on a windy day also illustrate the case in point.

The rapid evaporation of liquids in "vacuum" pans due to diminished pressure is an application of the principle that the rate of evaporation is increased by decreasing the pressure.

249. Heat of Vaporization. If heat be applied to water, the temperature rises until it reaches the boiling point. Further application of heat produces no further rise of temperature, the additional heat being used in changing the water from the liquid to the vapor state. *Heat of vaporization* is the heat required to change unit mass of a substance from the liquid to the vapor state *without changing its temperature.* The heat which is required to change one gram of water at 100° to steam at the same temperature is 539 calories; therefore we say that *the heat of vaporization of water at* 100° *C is* 539 *calories*

per gram, or in English units, 970 *B.t.u. per pound.*

The high heat of vaporization of water explains the value of steam heat as a means of heating buildings. Every gram of steam that changes from steam at 100° C to water at the same temperature gives up 539 calories of heat.

250. Vapor Pressure and Vapor Tension Compared. When a liquid evaporates into the space above it, the vapor thus formed exerts a pressure. *Vapor pressure* is the pressure which is exerted by an unsaturated vapor. *Vapor tension,* on the other hand, is the pressure exerted by a vapor which is saturated, that is, a vapor which is in contact with its liquid. Each liquid has its own definite vapor tension for a given temperature. If a tube 80 centimeters in length be filled with mercury and inverted in a dish of mercury, Fig. 210, a Torricellian vacuum, *v,* results. Now introduce into this tube at the bottom, by means of a glass pipette, a few drops of water which being lighter than the mercury rise in the tube to the surface, and a portion of it evaporates. The water vapor thus liberated in the space above the mercury exerts a pressure which forces the column of mercury down to a certain point *b*. The difference between the height of the column *a* and that of *b* represents in centimeters the vapor tension of water at the given temperature. Now if the tube be forced down into the mercury as shown in *c,* some of the vapor will liquefy, but the mercurial column will stand at the same level as before. The vapor tension of a liquid is the same whether the space above the mercury be great or small, *provided there is some free liquid present.*

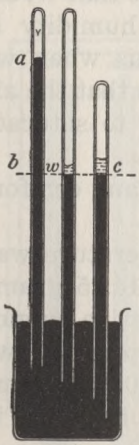


FIG. 210
Vapor pressure apparatus

251. Humidity. There is in the air at all times a certain amount of water vapor. At 20° C (68° F) a cubic meter of space when saturated contains about 17 grams of water vapor. The presence of air in the space does not appear to affect the amount of water vapor which it can contain.

Humidity is a term which is used to indicate the quantity of water vapor in the air. When the air contains all the water vapor it can at a given temperature, it is said to be *saturated.* At 20° C, for example, air is saturated when it contains 17.118 grams of water vapor per cubic meter, Table XIV, Appendix.

Relative humidity is the ratio of the amount of water vapor in the air at a given time to the amount that would be present if it were saturated. Relative humidity is usually expressed in terms of per cent; thus when we say that the relative humidity is 50, we mean that the air contains 50 per cent of the water required to saturate it at that temperature.

The relative humidity required for health and comfort in rooms is about 60 per cent.

Example. On a given day when the temperature was 23° C, it was found that the air contained 14.25 grams of water vapor per cubic meter. Find the relative humidity. *Solution*: On consulting Table XIV, Appendix, we find that air at 23° C is capable of holding 20.355 grams of water vapor per cubic meter. The relative humidity is therefore $14.25/20.355 = 70$ per cent.

EXERCISE. **17.** A given quantity of air at 20° C contains 10.27 grams of water vapor per cubic meter. What is the relative humidity?

252. Dew Point. Place a thermometer in a test tube partly filled with ether, Fig. 211. By means of a bent glass tube blow gently through the liquid a stream of air bubbles which will cause the ether to evaporate rap-

idly, thus producing a fall in temperature, as indicated by the thermometer. As the temperature falls, a point will be reached at which moisture or dew, condensed from the surrounding air, will appear on the outside of the tube. The temperature at which this moisture appears is called the dew point. In other words, the *dew point* is the temperature at which the water vapor of the air reaches saturation and begins to condense. Suppose, for example, that the air on a given day when the temperature is 20° C contains 12.7 grams of water vapor per cubic meter. Now it has been found by experiment (Table XIV) that the saturation temperature for 12.7 grams per cubic meter is 15° C. This means that if the temperature of the air falls to 15° C, water will be precipitated in the form of dew. In other words, for 12.7 grams of water vapor per cubic meter, 15° C marks the dew point. Fogs, clouds, and rain result from a lowering of the temperature below the dew point.

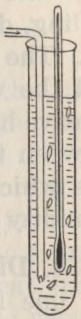


FIG. 211
Dew point apparatus

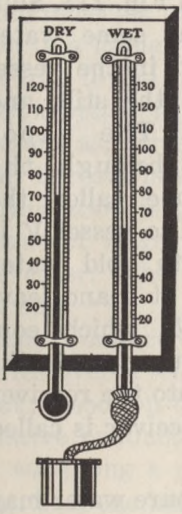


FIG. 212
Wet-dry bulb hygrometer

253. **The Hygrometer.** A *hygrometer* is an instrument for measuring relative humidity. There are several different kinds of hygrometers. The one shown in Fig. 212 is known as the wet-and-dry-bulb type. It consists of two thermometers placed side by side. The bulb of one is enclosed in wick, the lower end of which is inserted in a small vessel of water. The water rises in the wick by capillary action and evaporates, producing a cooling ef-

fect. Evaporation always tends to lower the temperature. The rate of evaporation, and consequently the rate of cooling, depends upon the amount of water vapor in the air. The difference between the readings of the dry-bulb and the wet-bulb thermometer will thus depend on the relative humidity of the air. By comparing the difference between the readings of the two thermometers with tables which accompany the instrument, both the relative humidity and the dew point can be found.

254. Distillation. *Distillation* is a process of separating one liquid from another or of separating a liquid from impurities by heating the mixture in one vessel and condensing the vapor in another. One form of distilling apparatus is shown in Fig. 213. Suppose that some water be placed in the vessel *S*, called the still, and heated. The vapor passes through the coiled tube called the worm. The vessel *V* is filled with cold water entering at *a* and leaving at *b*, which condenses the vapor in the coil *c*, from which the pure water drops into the receiver *R*. The liquid which is collected in this receiver is called the distillate.

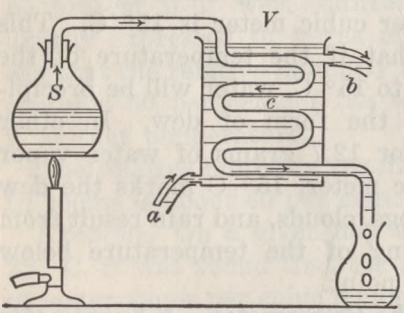


FIG. 213.—Distilling apparatus

By means of this process of distillation, pure water may be obtained from impure, and also one liquid may be separated from another, as for example, commercial alcohol from water, in which case alcohol being the more volatile, distills over more readily than the water. Mer-

cury is often distilled in order to separate it from its impurities.

The Transmission of Heat

255. Introductory. We wish now to consider the subjects of keeping foods cool and of keeping buildings warm and comfortable, in other words, the subjects of refrigeration and of heating and ventilation. In order to do this, however, we must first study the subjects of conduction, convection, and radiation, that is, the transmission of heat from one body to another. Heat may be transmitted from one point to another in three ways: (a) by conduction, (b) by convection, and (c) by radiation.

Conduction is the transmission of heat through a body from molecule to molecule, as in the heating of a piece of iron in a flame. The heat is transmitted from one end of the iron to the other by conduction.

Convection is the transmission of heat by means of a current as, for example, winds and ocean currents. The Gulf Stream is an example of a convection current.

Radiation is the transfer of energy through space by means of waves set up in a hypothetical medium called the *ether*. The earth is thus heated by radiation from the sun.

256. Conductivity of Solids. The conductivity of solids varies over a wide range. Metals are good conductors of heat; wood, on the other hand, is a poor conductor. The relative conductivity of metal and wood may be shown by wrapping a piece of paper tightly around a cylinder, Fig. 214, one half of which is composed of metal and the other of wood. If now the cylinder be held in a flame, the paper will char where it touches the wood. It will not be burned, however, where it is in contact with the metal. due to the fact that the metal conducts the heat

away so rapidly as to keep the paper below the burning point.

The conductivity of any given substance is expressed in the number of heat units conveyed in unit time through unit cross sectional area and unit thickness per degree of difference of temperature on the two sides. The conductivity of silver is usually taken as 1, which means that 1 calorie of heat will pass through 1 square centimeter of a silver plate which is 1 centimeter thick in 1 second when the difference of temperature on the two sides of the plate is 1° C. The relative conductivities of a few substances (including air) are given in the following table, a more complete list being given in Table XV, Appendix.

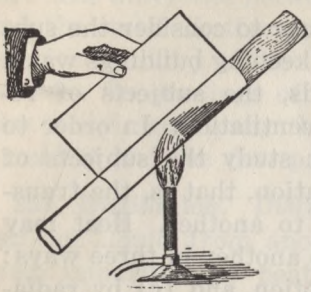


FIG. 214.—Metals are good conductors; wood is a poor conductor

THERMAL CONDUCTIVITIES

Silver	1.00	Wood	0.00040
Aluminum	0.50	Sawdust	0.00012
Iron	0.15	Air	0.00005

It thus appears that silver is twice as good a conductor as aluminum, and aluminum is more than three times as good a conductor as iron. Still air, as when confined in the interstices of wood, felt, or fiber, is almost a perfect nonconductor. The conductivity of all woody fibrous substances is extremely low. Much attention has been given of late to the preparation of fibrous heat insulating materials from waste products such as sawdust, straw, corn stalks, and sugar cane fibre. This material is manufactured under various trade names and is put out in board-like sheets or in granular form. The space between the

inner and outer walls of steel railway coaches is packed with this granular nonconducting material. Sheets of asbestos, because of its nonconducting and fire-proof properties, is extensively used to cover furnace and steam pipes.

The quantity of heat conducted through a given substance in a given time may be expressed in equational form as

$$H = \frac{K \times A \times T \times (t - t')}{L} = calories$$

in which K = coefficient of conductivity, A = area in cm^2, T = time in seconds, $(t - t')$ = difference in temperature in degrees C, and L = thickness in cm.

Example. A square tank 1 m on each edge is filled with water which is kept at a constant temperature of 70° C in a room the temperature of which is 20° C. The material of which the tank is made is copper, 1 mm in thickness. How much heat will be conducted through one side of this tank in half an hour? *Solution*: Consulting Table XV, Appendix, we find that the coefficient of conductivity of copper is 0.9. The conducting area $A = 100 \times 100$ cm^2, time $T = 30 \times 60$ seconds, difference of temperature $(t - t') = 70 - 20 = 50°$ C, and thickness $L = 0.1$ cm. Then $H = 0.9 \times 10000 \times 1800 \times 50/0.1 = 81 \times 10^8$ calories.

257. Conductivity of Liquids. Liquids are poor conductors of heat, the conductivity of water, for example, being about one one-thousandth that of silver, which is sometimes taken as a standard for the conductivity of metals.

Fasten a piece of ice in the bottom of a test tube nearly full of water. Hold the upper part of the tube in a flame, Fig. 215. In a short time the water in the

top of the tube will begin to boil, thus giving boiling water, tepid water, and ice water all in the same vessel.



FIG. 215.—Liquids are poor conductors of heat

The student should make note of the fact that when we speak of the conductivity of liquids, we are considering them as in a state of rest, that is, free from currents. When currents are set up in liquids, we have a case of conduction by *convection*, which will be studied later on.

258. Conductivity of Gases. Gases are extremely poor conductors of heat. Woolen clothing owes its nonconducting properties partly to the fact that the wool fibre is a poor conductor, but more largely because of the nonconducting property of the air enclosed within it. The fur of animals is an effective protection from cold for the same reason. The air spaces between double windows and between the walls of houses are also illustrations of some of the uses made of the nonconducting properties of gases.

The *fireless cooker* owes its heat retaining properties mainly to the air which is enclosed in the spaces of the material with which it is packed. This device, Fig. 216, consists usually of a wooden box containing a number of receptacles designed to receive the food to be cooked. Around these vessels there is closely packed some material, such as hay, felt, asbestos, or other nonconducting material, which encloses within its interstices nonconducting air. The food to be cooked is heated to the boiling point and then placed

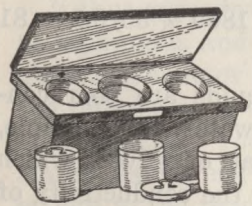


FIG. 216
Fireless cooker

in the receptacles within the cooker and carefully covered. Due to the nonconducting property of the packing material, and especially to the air enclosed within its spaces, sufficient heat is retained in the food until it is cooked. The efficiency of the apparatus depends, of course, upon the retention of the heat imparted to the food before it is placed in the cooker.

The *thermos bottle,* Fig. 217, differs from the fireless cooker in that it is designed to keep heat in or out, as desired. It consists of two bottles blown one inside the other and sealed together. The outside of the inner bottle, and the inside of the outer bottle are silvered. The air is exhausted from the space between the bottles, and this space is sealed airtight. It will be remembered that there are three ways in which heat is transmitted from one place to another, namely, by conduction, by convection, and by radiation. It is difficult for heat to enter or leave the bottle in any of these three ways. The evacuated space between the walls is an almost perfect nonconductor of heat. Also since the bottle is tightly closed no convection currents can occur. And finally, heat energy cannot be transmitted through the sides of the bottle by radiation because when the heat waves reach the bottle they are reflected by the bright silvered surface in the same way that light is reflected by the surface of a mirror.

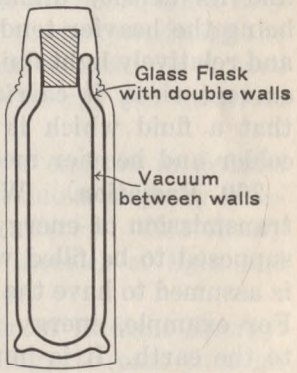


FIG. 217.—Thermos bottle

259. Convection. *Convection* is the transmission of heat by currents. A good illustration of convection currents is given in Fig. 218, *I* and *II*. In Fig. *I* the convec-

tion current rises from a hot portion of the beaker, the colder and therefore denser water at *CC* forcing it upward. In Fig. *II* the reverse motion of the convection current is depicted.

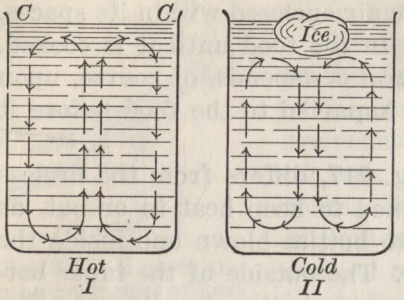


FIG. 218.—Convection currents

An explanation of convection may be given somewhat as follows: When a given mass of a fluid (air or water) is heated, it expands and its density diminishes. The surrounding medium being the heavier tends, therefore, to displace the warm and relatively light fluid, forcing it upward. As the warm current rises, it carries heat with it. Thus it appears that a fluid which is heated is forced upward by the colder and heavier medium surrounding it.

260. Radiation. We have defined *radiation* as the transmission of energy by means of waves. All space is supposed to be filled with a medium called ether, which is assumed to have the properties of transmitting energy. For example, energy is thus transmitted from the sun to the earth. It is important to note in this connection that in the transmission of energy from the sun to the earth, the intervening space is not heated, and therefore we cannot properly speak of this radiant energy as heat. It is true that the waves or impulses by means of which energy is transmitted may have their origin in a heated body, the sun, and when they fall upon the earth they may give rise to heat, which is transmitted from point to point through the medium of matter by conduction or convection.

It must not be thought, however, that the sun is the

only source of radiant energy. According to the emission theory, all matter is capable of emitting energy at all times and under all conditions. It was assumed by a German physicist named Max Planck (pronounced Plänk) that this energy is not given off in a continuous stream, but rather by a series of jerks or impulses called quanta (plural of quantum). Thus, when an atom emits energy by radiation the energy does not flow out like a stream of water from a hose but rather like a stream of bullets from a machine gun, each bullet representing a quantum. A *quantum* is sometimes called the "atom of energy." The rate of emission of quanta from the surface of a body depends upon the character of the surface and upon the temperature. And further, quanta vary in size, depending upon the color of the body emitting the radiation. For example the quanta emitted by a blue body contains about twice as much energy per quantum as those emitted by a red body.

The theory that radiant energy is emitted in definite energy units or quanta in a discontinuous or jerky manner is known as *Planck's quantum theory.* This theory has important applications to the theories of light and electricity, as will be explained later.

261. Transmission and Absorption of Radiant Energy. Some substances allow radiant energy to pass through readily. Rock salt transmits radiant energy the most readily of any substances known. Other substances do not transmit radiant energy readily but absorb it. Lampblack absorbs radiant energy to a greater degree than any other known substance.

Radiant energy, as has been stated, is assumed to be transmitted by means of ether waves, some of which are long and some are short. Some substances like glass and water allow the short waves to pass through readily, but absorb the long waves. Radiant energy comes from

the sun to the earth in the form of both long waves and short waves. Now the atmosphere surrounding the earth contains enormous quantities of water vapor, which transmits the short waves and shuts off the long waves. These short waves on striking the earth heat it, and give rise to long waves, which are prevented by the water vapor in the atmosphere from passing out into space. Thus the water of the atmosphere serves as a sort of protecting blanket, which by shutting out the long waves keeps our days from being excessively hot, and by preventing too rapid radiation, keeps our nights from being excessively cool.

The hot house has been called, and quite rightly, a trap to catch sunbeams. The short waves of radiant energy pass readily through the glass and falling upon the interior heat it. Now the heat thus generated gives rise to long waves which cannot pass through the glass, their energy being thus entrapped, so to speak. That glass allows short waves to pass and not long waves may be seen when we consider the fact that the short waves conveying radiant energy from the sun pass readily through the glass of windows, and falling on our bodies give rise to heat. On the other hand, it is a well known fact that if a glass screen be placed in front of a grate fire it will effectively shut off the heat; that is, it will prevent the transmission of the long waves conveying radiant energy from the fire.

262. Illustration of the Absorbing Power of Lampblack. To show that lampblack absorbs radiant energy to a greater degree than does glass, take two air thermometers as shown in Fig. 219 and coat the bulb of one of them with lampblack from a candle flame. Adjust the colored liquid to about the same height in both stems. Now place a hot body, such as a Bunsen flame or a sheet of hot iron midway between the bulbs. The radiation

from the hot body passes quite readily through the glass bulb, but is absorbed by the blackened bulb, as is shown by the lowering of the colored liquid in the tube.

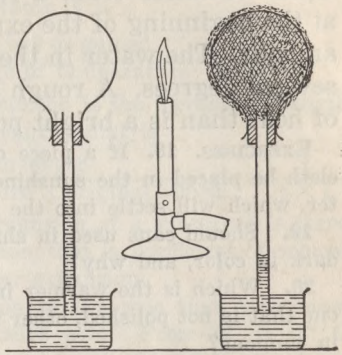


FIG. 219.—Illustrating the absorbing power of lampblack

263. Absorption and Reflection of Radiant Energy. The amount of radiant energy absorbed or reflected by a body depends upon (a) the temperature of the body and (b) the nature of its surface. In general a rough black body is a good absorber but a poor reflector of radiant energy; a smooth bright body, on the other hand, is a poor absorber but a good reflector.

To determine the relative radiating power of a black body as compared with that of a bright body, take two bright tin cans (baking powder cans from which the wrappings have been removed will do) and punch a hole in the cover of each for the admission of the thermometers, Fig. 220. Coat one of the cans heavily with lampblack. Fill both with hot water and note the temperature of each at the beginning of the experiment. After a time again observe the temperature of the water in each can. It will be found that the water in the black can is considerably cooler than that in the bright one. A rough black body radiates more readily than a smooth bright body.

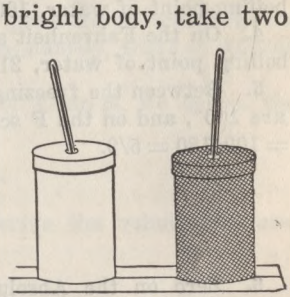


FIG. 220.—A bright body is a good reflector and a poor absorber of radiant energy

Again fill both cans with cold water and insert the

thermometers. Place the cans in the sunshine, or at equal distances from a radiator. Note the temperature at the beginning of the experiment and at the end of half an hour. The water in the black can will be the hotter by several degrees. A rough black body is a better absorber of heat than is a bright polished body.

EXERCISES. **18.** If a piece of black cloth and a piece of white cloth be placed in the sunshine upon the surface of snow in winter, which will settle into the snow the more rapidly? Explain.

19. Should cans used in shipping milk in summer be bright or dark in color, and why?

20. Which is the warmer in winter, a shoe that is polished or one that is not polished, other things being equal? Which is cooler in summer?

SUMMARY

1. Heat is a form of energy due to the molecular motions of a body.

2. Temperature is a measure of the degree of hotness or coldness of a body.

3. On the Centigrade scale the melting point of ice = 0°; the boiling point of water, 100°.

4. On the Fahrenheit scale the melting point of ice = 32°; the boiling point of water, 212°.

5. Between the freezing and boiling points on the C scale there are 100°, and on the F scale 180°; the ratio of C units to F units $= 100/180 = 5/9$.

$$C = \frac{5}{9}(F - 32), \text{ and}$$

$$F = (\frac{9}{5}C) + 32$$

6. Zero on the Absolute scale is the point of no molecular motion, that is, 273 degrees below 0 on the C scale. Hence C + 273 = Absolute.

7. Unit quantity of heat: The calorie is the heat required to raise the temperature of 1 gram of water 1° C; the B.t.u. is the quantity of heat required to raise the temperature of 1 pound of water 1° F.

8. Specific heat is numerically equal to the number of heat units required to change the temperature of unit mass one degree; that is, specific heat of a substance is the number of calories re-

quired to raise the temperature of 1 gram 1° C, or, in English units, the number of B.t.u. required to raise the temperature of 1 pound 1° F. The specific heat of water is 1.

9. $\text{Specific heat} = \dfrac{\text{heat units lost or gained}}{\text{mass} \times \text{change of temperature}} = \dfrac{H}{m \times t}$

10. The coefficient of linear expansion is the increase of length of a body per degree per unit of length; that is

$$\text{L. C. E.} = \frac{\text{change in length}}{\text{change in temperature} \times \text{original length}} = \frac{l}{t \times L}$$

11. The coefficient of cubical expansion of a substance is three times its coefficient of linear expansion, that is, coefficient of cubical expansion = 3 × coefficient of linear expansion.

12. Water is a partial exception to the rule that substances expand with heat and contract with cold. From 0° to 4° C water contracts, after which it expands with rise of temperature. Water has its maximum density at 4° C.

13. All gases have the same coefficient of expansion, namely, 1/273 or 0.00366 of the volume at 0° C.

14. Charles' Law: When the volume is constant, the pressure coefficient for all gases is the same, namely, 1/273 = 0.00366 of the pressure at 0° C.

15. Gay-Lussac's Law: When the pressure is constant, the volume coefficient of all gases is the same, namely, 1/273 = 0.00366 of the volume at 0° C.

16. As a deduction from the law of Charles, we derive the pressure-absolute-temperature equation

$$\frac{P}{P'} = \frac{T}{T'}$$

17. From Gay-Lussac's law, we derive the volume-absolute-temperature equation

$$\frac{V}{V'} = \frac{T}{T'}$$

18. The standard hydrogen thermometer is an instrument for measuring temperature in terms of pressure.

19. Fusion or melting: The melting point of a substance is the temperature at which it changes from the solid to the liquid state. Only crystalline substances have definite melting points; non-crystalline substances, such as wrought iron and glass, change gradually from the solid to the liquid state. The melting point of ice is 0° C or 32° F.

20. The heat of fusion of a crystalline substance is the number of heat units required to change unit mass of the solid to a liquid, without change of temperature. The heat of fusion of ice is 80 calories per gram or 144 B.t.u. per pound.

21. Summary of the principal facts relating to the fusion or melting of crystalline solids: (a) For crystalline substances, the freezing point and the melting point are the same. (b) Every crystalline substance has a definite melting point for a given pressure. (c) When a crystalline substance reaches its melting point, its temperature remains constant until it is entirely melted.

22. Boiling: The boiling point of a liquid is the temperature at which bubbles of vapor form at the bottom of the containing vessel and rise to the top. When the boiling point is reached, an increase of heat does not change the temperature of the liquid; it only increases the rate at which the liquid is boiled away.

23. Summary of the main facts relating to boiling: (a) Every liquid has a definite boiling point which is invariable under the same conditions. (b) An increase of pressure raises the boiling point; a decrease of pressure lowers the boiling point. (c) Salts dissolved in a liquid raise the boiling point; gases dissolved in a liquid lower the boiling point.

24. Vaporization is the change of a substance to the vapor state; it may be by boiling, evaporation, or sublimation. When a substance changes from a solid to a vapor without passing through the liquid state, it is said to sublime.

25. Heat of vaporization is the number of units of heat required to change unit mass of a liquid at a given temperature to vapor at the same temperature. The heat of vaporization of water at 100° C is 539 calories per gram, or 970 B.t.u. per pound.

26. Evaporation is the process by which a liquid changes to the vapor state quietly at the surface. The main facts relating to evaporation are as follows: (a) The rate of evaporation increases with the free surface of the liquid. (b) The rate of evaporation increases with an increase of temperature. (c) The rate of evaporation increases with a change of air in contact with the liquid. (d) The rate of evaporation increases as the pressure diminishes.

27. Humidity refers to the quantity of water vapor in the air at any given time. When for a given temperature the air contains all the water vapor it can hold, it is said to be saturated.

28. Relative humidity is the ratio of the quantity of water

vapor in the air at any given time to the quantity required for saturation; that is

$$\text{relative humidity} = \frac{\text{water actually in the air}}{\text{quantity required for saturation}}$$

29. Dew point is the temperature at which water vapor in the air reaches saturation. Note that relative humidity is a *ratio*, while dew point is a *temperature*.

30. Transmission of heat may be by conduction, convection, or radiation.

31. Conduction is a transfer of heat from molecule to molecule. Of solids, metals are good conductors, while organic substances, such as wood, felt, and fiber, are poor conductors. Liquids are relatively poor conductors and gases very poor.

32. Thermal conductivity, $H = KAT\ (t - t')/L$ in which K = coefficient of thermal conductivity, A = area, T = time, $(t - t')$ = difference in temperature, and L = thickness of the conductor.

33. Convection is the transfer of heat by currents. Convection currents are made use of in the heating and ventilation of buildings.

34. Radiation is the transfer of energy by ether waves. Rough and blackened surfaces are good absorbers and good transmitters of radiant energy; bright polished surfaces are poor absorbers and poor transmitters of radiant energy.

REVIEW EXERCISES

A

1. Discuss briefly the nature of heat. Tell about the work of Count Rumford and Sir Humphrey Davy.

2. Define temperature, and tell why the sensations of hot and cold are unreliable means of measuring it.

3. Describe the mercury thermometer and the principle upon which it works.

4. How are the fixed points on a thermometer determined? What are the readings for the freezing and boiling points on the Centigrade scale? the Fahrenheit scale?

5. Explain by means of a sketch the relation of the C scale to the F scale for the temperature range included between the freezing point and the boiling point.

6. Give the equivalent F readings for the following: (a) 70° C; (b) 50° C; (c) 25° C; (d) —5° C; (e) —20° C.

7. Give the equivalent C readings for the following: (a) 113° F; (b) 77° F; (c) 14° F; (d) —13° F.

8. What is the difference between temperature and quantity of heat?

9. Explain the distinction between the calorie and the B.t.u.

10. What is the specific heat of a substance that requires 264 calories to raise the temperature of a mass of 100 g from 20° C to the boiling point of water? What is the substance?

11. How many calories of heat will be required to change the temperature of 100 grams of the following substances from 0° C to 100° C: (a) water? (b) iron? (c) lead?

12. Give examples of the practical applications of the expansion of solids.

13. (a) Define coefficient of linear expansion. (b) How may the coefficient of cubical expansion for a given substance be derived from its coefficient of linear expansion?

14. A brass rod at 0° C has a length of 200 cm; at 100° C its length is 200.36 cm. Find its coefficient of expansion.

15. What is the coefficient of expansion of gases, and wherein does it differ from the coefficient of expansion of solids?

16. Wherein does the law of Charles differ from that of Gay-Lussac? What factor is kept constant in each case?

17. Explain the meaning of each term in the following equations: (a) $P/P' = T/T'$; (b) $V/V' = T/T'$.

18. A given mass of gas has a volume of 250 cm^3 at 0° C. Find its volume at (a) +10° C; (b) —10° C, the pressure remaining constant.

19. A given mass of gas free to expand has a volume of 500 cm^3 at —13° C. What will be its volume at +27° C, the pressure remaining constant?

20. Explain how bodily comfort is dependent upon the humidity conditions of the atmosphere.

21. What is the distinction between humidity and relative humidity?

22. What is meant by the statement that the heat of fusion of ice is 80 calories?

23. State the principal facts relating to the fusion of crystalline solids.

24. How many calories of heat will be required to change 10 grams of ice at zero to water at (a) 0° C; (b) 20° C?

25. State two ways by which the boiling point of water may be raised.

26. State the principal facts relating to boiling, and explain their application to the boiling of water.

27. What is the distinction between vaporization and evaporation? Upon what factors does the rate of evaporation depend?

28. What is meant by the statement that the heat of vaporization of water at 100° C is 539 calories?

29. How much heat is required to change 100 grams of ice at 0° C to 100 grams of steam at 100° C?

30. One kilogram of steam in a radiator at 100° C condenses to water at 100° C. How many calories of heat are given out?

31. Describe the radiating power and the absorbing power of lampblack as compared with that of a metal having a bright, polished surface.

32. What useful purpose is served by water vapor in the atmosphere, so far as living conditions on the earth are concerned?

33. Why is light colored clothing cooler in summer than black clothing?

34. What is the objection to painting steam radiators with aluminum or "gold" paint?

B

35. Find the boiling point and the freezing point of mercury (a) on the C scale; (b) F scale.

36. Find the boiling and freezing points of alcohol on (a) the C scale; (b) F scale.

37. One hundred grams of mercury at 100° C are stirred with 100 grams of water at 20° C and the resulting temperature is 22.5° C. What is the specific heat of mercury?

38. Five hundred grams of copper, specific heat 0.09, temperature 120° C, are dropped into 500 grams of water at 20° C. What is the resulting temperature?

39. Explain how to find the specific heat of a body by the method of mixtures.

40. Explain each term in this equation: $m \times t \times s = m' \times t' \times s' + m'' \times t'' \times s''$. If water is used in the calorimeter, what is the numerical value of s'?

41. Discuss the expansion of water from 0° C to 8° C.

42. The coefficient of expansion of nickel steel is 0.00001. A

rod of this material having a length of 5 meters at 20° C will have what length in centimeters at 12° C?

43. At 10° C the air pressure in an automobile tire is 35 lb. per sq. in. What will be the pressure at 77° C, assuming that the tire does not expand appreciably?

44. Two pounds of steam in a radiator at 212° F condenses to water, the temperature of which falls to 190° F. How many B.t.u. are given out?

45. How many calories of heat will be required to change 10 grams of ice at —10° C to water at the boiling point, the specific heat of ice being 0.5?

46. To what elevation in feet above sea level must water be taken in order that its boiling point shall be lowered to 95° Centigrade?

C

47. A liter of gas at —3° C expands under constant pressure due to change of temperature, to 1200 cm^3. Find the change of temperature in C degrees.

48. On a certain winter's day when the temperature was 41° F, air was found to contain 6.22 g of water vapor per m^3. Express the relative humidity of this air in per cent.

49. Air, as described in Ex. 48, was heated and allowed to enter a living room at a temperature of 68° F, without the addition of moisture in passing through the furnace. (a) Find the relative humidity of the air in the room after being heated. (b) How much was this humidity below that required for health and comfort?

50. Compare the relative conductivities of solids, liquids, and gases.

51. In 5 minutes how many calories of heat will be conducted through the bottom of an aluminum stew pan, the area of which is 350 cm^2 and the thickness 1 mm, when the difference of temperature between the two sides is 120° C?

CHAPTER VIII

HEAT (Continued)

Refrigeration

264. Introductory. In modern life the problem of keeping food and buildings cool is almost as important as the problem of heating and ventilation. The preservation of food by the cold storage system, as well as the manufacture of artificial ice, involves an outlay of millions of dollars annually. The artificial cooling of buildings, such as hotels and auditoriums, is today receiving much attention by builders and designers.

There are two general methods of securing refrigeration, (a) by natural means, such as the use of ice, and (b) by the production of cold by artificial means. We shall first consider the method of producing cold by the use of ice. In order to cool food to about 50° F (10° C) and keep it at this temperature, heat must be absorbed. This is conveniently done by allowing the food to supply the heat to melt the ice. When ice is melted, the amount of heat absorbed is equal to 80 calories per gram or 144 B.t.u. per pound of ice consumed. If the price of ice is 50 cents per 100 pounds, the cost of absorbing heat is about 10 times the cost of producing heat by burning city gas at $1.50 per 1000 cubic feet. Unfortunately the melting of ice cannot be shut off as can a gas flame. This fact makes it especially important to prevent as far as possible heat from getting to the ice. To do this we enclose the ice in a properly constructed refrigerator or ice box having well insulated walls.

265. The Ice Box. The inner chamber of an ice box is

usually made of galvanized or enameled metal. The outer casing is made of wood. The walls of a good ice box usually contain two or more nonconducting solid or semi-solid media, Fig. 221. The fiber or felt packing is especially effective as a nonconductor because of the air spaces which it contains. The ice box contains a number of compartments, one for the ice and the others for food which it is desired to keep cool, Fig. 222. The ice is placed in a compartment near the top of the box. The air next the ice becomes cool and sinks through the openings in the ice chamber into the main part of the refrigerator, while the warmer air is forced upward where

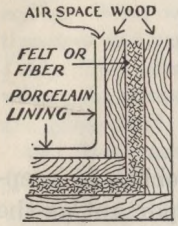


FIG. 221
Walls of an ice box

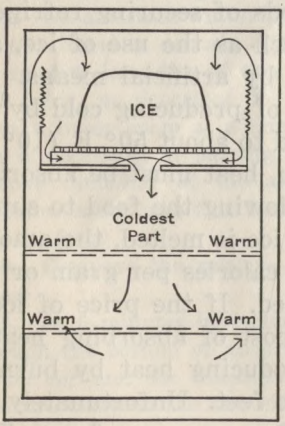


FIG. 222.—Diagram showing circulation of air in an ice box

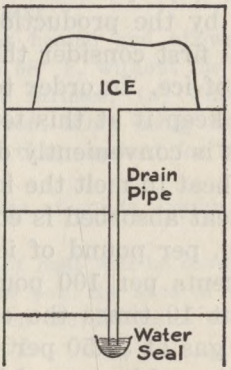


Fig. 223.—Water seal under ice box

it comes in contact with the ice and there becomes cool. There is thus a continuous circulation of air through the

ice chamber and down into the food compartment. This circulation is important because by this means the cooled air is distributed to all parts of the refrigerator, and also because on passing over the ice the air loses some of its moisture and consequently some of the odors which it has taken up from the food. It follows, therefore, that anything which retards this circulation interferes seriously with effective refrigeration.

Water from the melting ice is carried away through a vertical drain pipe which should be closed by means of a water seal or trap, Fig. 223. The water trap is a very important adjunct to the ice box because it allows the waste water to drain out and at the same time prevents a current of warm air from flowing upward into the ice chamber. When the water trap by chance becomes detached from the drain pipe, there results a rapid and wasteful melting of the ice.

266. Proper Care of the Ice Box. Recent experiments carried out on an extensive scale have revealed the fact that many ice boxes in households are not only wasteful of ice but in some cases dangerous to health because of faulty low temperature maintenance.

Investigation by the U. S. Department of Agriculture has shown that in milk kept at 60° F about 15 times as many bacteria will develop in a day as in milk kept at 50° F. It is important, therefore, that such foods as milk and meats should be kept in the coldest part of the ice box, which is the part immediately below the ice.

Slow melting of the ice does not necessarily indicate a good refrigerator. Unless the ice melts it can absorb no heat and it is therefore of no use as a cooling agent. Protecting the ice in a refrigerator by covering it with cloth or paper is a good way to save ice but a poor way to save food. The only proper way to economize in ice is to use a refrigerator with better insulating walls, and by

opening the doors as seldom and for as short a time as possible. In order that the efficiency of an ice box be high, the following conditions should be noted: (a) The insulation should be as perfect as possible; (b) the ice chamber should be kept full; (c) doors should be kept closed as much as possible; (d) the drainage tube should be kept closed by means of a water trap.

In many households at the present time the electric refrigerator has displaced the ice box, thus doing away entirely with the services of the ice man. The electric refrigerator will be discussed later on.

267. Production of Cold by Artificial Means. The production of cold by artificial means is today of great commercial importance. There are three methods of producing cold artificially, namely, (a) by solution, (b) by evaporation, (c) by expansion of gases. The fundamental principle involved in these three methods is the same; that is, each process involves the expenditure of energy, which in turn involves the absorption of heat. When, for example, salt dissolves in a liquid, or a liquid evaporates, or a gas expands, energy in the form of heat is taken and a condition of cold is produced.

268. Cold by Solution. When a lump of sugar is dropped into a cup of coffee, the temperature of the coffee is lowered slightly as the sugar dissolves, due to the fact that the dissolving of the sugar takes energy from the liquid in the form of heat. In general the solution of crystalline substances in liquids tends to lower the temperature, although in some cases this lowering of temperature is masked by secondary chemical reactions. If a mixture of sodium thiosulphate ($Na_2S_2O_3$), known as photographer's "hypo," is put into a beaker of water at room temperature and stirred gently with a thermometer, it will be observed that the temperature of the water will fall several degrees. The dissolving of the salt takes

energy from the water in the form of heat, thus causing the fall in temperature.

269. The Ice Cream Freezer. The common ice cream freezer is one of the best known devices for producing cold by artificial means. The principle employed is that of the production of cold by solution. The cream to be frozen is placed in a metallic vessel *A*, Fig. 224, in which a stirrer is rotated by means of a crank *C*, or in large machines by means of a motor. Around this vessel there is packed a mixture of salt and crushed ice, *B*. The contact of the salt with the ice tends to cause the latter to melt, and the resulting liquid dissolves the salt, both operations absorbing heat, which is withdrawn from the cream in the vessel *A*. The object of stirring the cream is to bring all portions of it successively in contact with the walls of the containing vessel, and to break up the ice crystals which form. It is important to note that the cream freezes when the ice melts; that is, the melting ice absorbs heat, which is furnished by the cream.

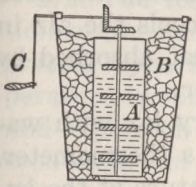


FIG. 224
Ice cream freezer

It has been found by experiment that the best mixture of ice and salt for producing a maximum lowering of temperature is 3 parts of ice to 1 part of salt by weight.

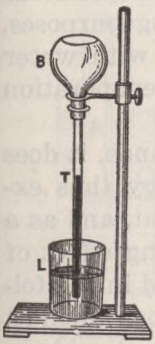


FIG. 225
Air thermometer

270. Cooling by Evaporation. When a liquid evaporates, energy in the form of heat is required to separate the molecules. Evaporation is, therefore, a cooling process. A gram of water at 20° C (ordinary room temperature) in changing from a liquid to a vapor takes up 585 calories of heat. If some highly volatile liquid, as ether, be dropped

upon the bulb of an air thermometer, Fig. 225, the colored liquid in the stem will quickly rise, due to the chill produced by the rapid evaporation of the ether, resulting in a contraction of air within the bulb. The student should not confuse liquid ether, which is a chemical compound, with the hypothetical "ether" which is supposed to fill all space. Sprinkling the floor with water in the summer cools the air in the room because of the heat absorbed by the evaporation of the water.

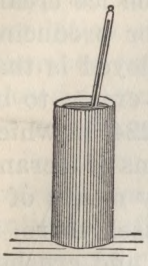


FIG. 226
Evaporation through porous jar producing a cooling effect

Fill a small porous battery jar with water and place in the vessel a thermometer, Fig. 226, noting the temperature at the beginning of the experiment. In the course of fifteen minutes the temperature of the water will have fallen considerably, due to evaporation from the sides of the jar. In some warm countries, as Mexico, water is often cooled by placing it in large porous vessels, which allow evaporation to take place over the entire surface. Also, Boy Scouts, and others, in camping expeditions frequently use a canvas water-bag for cooling purposes. This is a porous canvas bag, which is filled with water and hung up in a convenient place so that evaporation from the surface may take place.

271. Cold by Expansion. When a gas expands, it does work on the surrounding medium; the energy thus expended comes from the gas in the form of heat, and as a result the temperature is lowered. The cooling effect of an expanding gas may be strikingly illustrated in the following manner. Obtain a tank of liquid carbon dioxide, such as is used in soda fountains. The carbon dioxide in the tank is under high pressure. If now the stop cock connected with the tank be opened and the gas within allowed to expand, there will result a lowering of the

temperature, due both to the vaporization of the liquid and also to the expansion of the gas. The chill produced is sufficient to freeze the carbon dioxide, which appears in the form of a white mist or snow, Fig 227.



FIG. 227
Carbon dioxide snow

The temperature of solid carbon dioxide at atmospheric pressure is —78° C.

272. The Ice Machine. In the operation of the modern commercial refrigerating apparatus a reduction of temperature is obtained by the evaporation and expansion of liquid ammonia. Gaseous ammonia is especially suitable for this purpose because it may be liquefied by compression alone at ordinary temperatures. At 27° C (80° F) a pressure of 155 pounds per square inch is required.

The principle of the ice machine may be illustrated by means of a very simple appartus shown in outline in Fig. 228. A cylinder *AB*, which is provided with a piston and

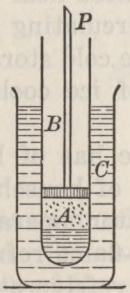


FIG. 228

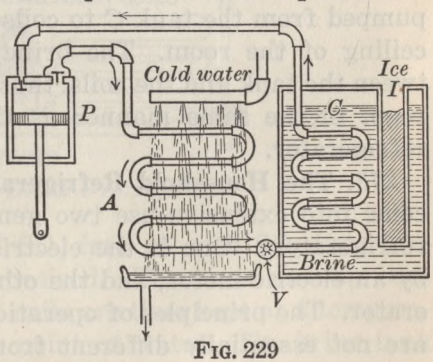


FIG. 229
Ice machine

which contains some liquid ammonia under high pressure, is placed in the water to be frozen. The piston is suddenly drawn up, thus reducing the pressure upon the liquid.

The ammonia in *A* evaporates and expands into *B*, producing a chill, and thus tending to freeze the water in *C*. The lowering of the temperature is produced by both the evaporation and the expansion of the ammonia. In actual practice the machine is very much more complicated, as shown in Fig. 229.

The pressure pump *P* is provided with two valves, one opening upward and the other downward. Liquid ammonia is placed under high pressure in the coils *A*, which are kept cool by means of a stream of cold water playing over them. The liquid ammonia is then allowed to expand through the valve *V* into the cooling tank *C*, which contains brine. As the ammonia expands (changes from a liquid to a gas) in the tubes of chamber *C*, heat is absorbed from the brine, reducing its temperature to about $-10°$ C. Ice cans *I* containing pure water are suspended in the cold brine. The water in the cans freezes, forming solid blocks.

When it is desired to cool a storage room in a house, hotel, steamship, or cold storage warehouse, the brine is pumped from the tank *C* to coils of pipe placed near the ceiling of the room. The brine is kept circulating between the tank and the coils, thus cooling the cold storage room in the same manner as the block of ice cools a refrigerator.

273. The Household Refrigerator. There has of late come into extensive use two general types of household refrigerators. One is the electric refrigerator, operated by an electric motor, and the other the gas-flame refrigerator. The principles of operation of these refrigerators are not essentially different from those involved in the operation of the commercial ice machine (Art. 272).

The general lay out of an electric refrigerating system is shown in outline in Fig. 230. The expansion substance employed is liquid sulphur dioxide (SO_2), which is highly

volatile. The pressure pump is run by a small electric motor, power being taken directly from the house lighting circuit. Liquid SO_2, which is forced under high pressure by the pump into the compression coils on the right, expands through the expansion valve at the top into the evaporating coils, which are immersed in brine. It is this

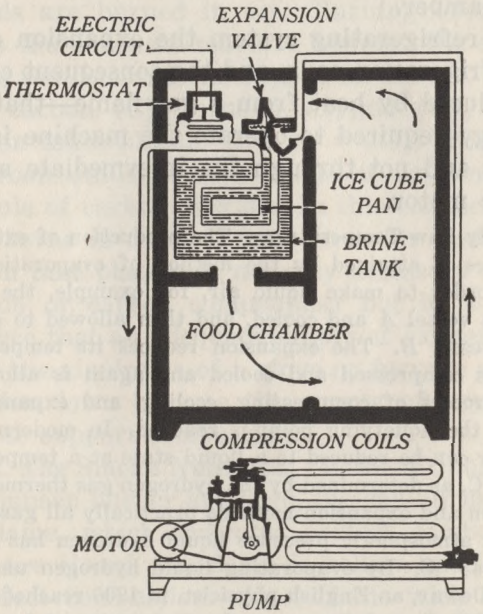


FIG. 230.—Electric refrigerator

expansion of the liquid SO_2 into gaseous SO_2 that produces the cooling effect desired. The brine and the contents of the evaporating coils are thus kept at a temperature below the freezing point of water. The brine tank and the evaporating coils take the place of the cake of ice of the ordinary ice box, only it is more efficient in keeping a constant temperature within the refrigerator. Constant temperature is maintained by means of a thermostat

which by expansion and contraction automatically closes and opens the electric circuit. Thus when the temperature falls below a fixed point, the circuit is broken and the motor stops; when the temperature rises to the freezing point of water, the circuit is closed and the motor starts. In this way an even, low temperature is maintained in the cooling chamber.

In the gas refrigerating system the expansion of the gas in the refrigeration coils, and the consequent cooling effect, is produced by heat from a gas flame—that is to say, the energy required to operate the machine is supplied by heat and not through the intermediate agency of an electric motor.

274. Extremely Low Temperatures. The production of extremely low temperatures is attained by the method of evaporation and expansion. In order to make liquid air, for example, the air is compressed in a vessel *A* and cooled, and then allowed to expand into a second vessel *B*. The expansion reduces its temperature. This cold air is compressed and cooled and again is allowed to expand. The process of compressing, cooling, and expanding is continued until the liquefying point is reached. In modern liquid air machines air can be reduced to a liquid state at a temperature of about —180° C, as determined by the hydrogen gas thermometer.

By evaporation and expansion methods practically all gases may be liquefied. At atmospheric pressure liquid hydrogen has a temperature of —243° C. By evaporating liquid hydrogen under reduced pressure Dewar, an English physicist, in 1900 reached a temperature of —260° C, and Onnes, at the University of Leyden in Holland, attained a temperature of —272.18° C, which is less than 1° C above Absolute zero. At these low temperatures steel becomes almost as brittle as glass, and in general the electrical conductivity of all metals is very much increased. Temperatures below the liquefying point of hydrogen have to be determined by electrical conductivity methods.

Heating Systems

275. Introductory. Fuels may be classified roughly as solid fuels, liquid fuels, and gaseous fuels. The solid

fuels in common use today are wood, charcoal, coal, and coke; liquid fuels: petroleum, refined oils, such as kerosene and gasoline, and alcohol; gaseous fuels: natural gas, manufactured gas, and acetylene.

The heat generated by an electric current will be discussed in the chapter on Electricity.

Fuels are burned in air. Burning or combustion is due to the chemical reaction between the elements composing the fuel and the oxygen of the air. For example, when carbon (C) burns in oxygen (O), the following reaction takes place: $C + 2O = CO_2$; that is, one atom of carbon unites with 2 atoms of oxygen to form one molecule of carbon dioxide. In this connection the point to be borne in mind is this: In order that perfect combustion take place it is not only necessary to supply fuel (carbon or carbon compounds) but it is of equal importance that an adequate supply of oxygen be available. When the air drafts of a stove or furnace are closed, the fire dies down because the supply of oxygen is inadequate to support combustion.

276. The Smoke Nuisance. If the air holes at the base of a Bunsen burner are open, the gas burns with a colorless flame, which indicates perfect combustion; if the holes are closed, the flame is yellow and smoky, due to imperfect combustion. Also, if the wick of a kerosene lamp be imperfectly trimmed or turned up too high, the lamp smokes; this is an excellent illustration of imperfect combustion. Smoke is always an indication of imperfect combustion; it also indicates a waste of fuel, because smoke consists mainly of particles of unburned carbon.

Smoke not only represents a waste of fuel amounting annually to millions of dollars, but in many of our cities it has become a nuisance so serious that stringent laws have been enacted looking toward its abatement. Pittsburgh furnishes an example not only of the smoke evil

at its worst, but also of what may be done to mitigate it. Exhaustive studies of the subject were made in that city



FIG. 231.—A Pittsburgh steel plant before the installation of smoke consuming and fuel saving devices

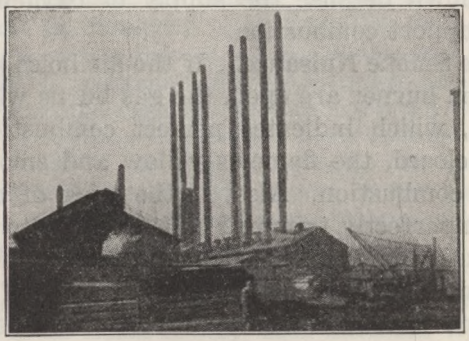


FIG. 232.—The same in full operation after the installation of smoke consuming devices

by the Mellon Institute of Industrial Research. Measurements of sootfall made at twelve stations in Pittsburgh

indicated an annual average deposit of soot in that city of more than 1000 tons per square mile. In Leeds, a manufacturing town in England, the sootfall ranges from 25 to 500 tons per square mile in different sections of the town. London's average is 248 tons per square mile for the whole city. It was found that the smoke evil was then (1912-13) costing Pittsburgh about $10,000,000 per annum, not including depreciation in the value of property, absence of various industries which cannot flourish in a smoky atmosphere, and injury to human health. As a result of the collection and publication of these data, stringent smoke-abatement ordinances were adopted in 1914, and in the following year it was reported that the amount of smoke emitted had been reduced 75 per cent.

The first smoke abatement invention was patented by James Watt in 1785. Present day methods adopted to diminish smoke output include (a) changes from soft coal to gas, coke, and low volatile fuels; (b) installation of down-draft boilers; (c) steam jets and extension stacks; and (d) automatic stokers. An automatic stoker is a mechanical device for delivering the fuel to the furnace at a uniform rate and for distributing it evenly upon the grates. This favors complete combustion, thus increasing the heat energy derived from the fuel and at the same time decreasing the smoke output.

In Fig. 231 we have a view of the smoke issuing from the furnace of a steel plant in the Pittsburgh district before the installation of smoke consuming and fuel saving devices. The same plant as it looks today is shown in Fig. 232. Both plants are in full operation.

277. Heat of Combustion. The *heat of combustion* of a substance is measured by the number of units of heat (calories or B.t.u.) given out when unit quantity of the substance is burned. For example, the heat of combustion of anthracite coal is 8000 calories per gram, or 13,000

B.t.u. per pound. This means that a ton of hard coal when completely burned will furnish $2000 \times 13{,}000 = 26{,}000{,}000$ B.t.u. of heat.

Fuels may be classified as solid, liquid, and gaseous. The solid fuels in use in this country are wood, charcoal, coal, and coke. Charcoal is made from the distillation of wood, and coke from the distillation of soft coal. Both burn with a practically smokeless flame and give out much heat, the heat of combustion being especially high.

Coal is classified with reference to the volatile matter that it contains as bituminous or soft coal, semi-bituminous, and anthracite or hard coal. (a) A bituminous coal is one that contains over 20 per cent of volatile matter. This coal is used largely in the manufacture of gas. Hocking Valley coal, from Ohio, is a type of bituminous coal of average grade, having a heat of combustion of about 14,000 B.t.u. per pound. (b) Semi-bituminous coal looks something like anthracite, but is lighter in weight and burns more readily. It is valuable where it is necessary to keep an intense heat. A high grade semi-bituminous, from West Virginia, is that known as Pocahontas, the heat of combustion of which is 15,000 B.t.u. per pound. (c) Anthracite coal, found only in eastern Pennsylvania, contains less than 10 per cent of volatile matter. It ignites very slowly and burns at a high temperature. Owing to its smokeless burning it is used almost altogether for domestic purposes. A good quality of anthracite coal, such as Scranton coal, has a heat of combustion of about 13,000 B.t.u. per pound.

The liquid fuels are gasoline, kerosene, and crude oils. In recent years crude oils are being used in increasing quantities as a source of heat for the engines of railway trains and ocean going vessels. They are now being used quite extensively also in the modern automatic oil burners for the heating of dwelling houses. The heating power

of kerosene is about 20,000 B.t.u. per pound and that of gasoline 20,250 B.t.u.

The gas fuels comprise natural gas and manufactured gas, commonly known as illuminating gas. In the economical use of gas as a fuel for cooking, gas stove burners should be so adjusted that the blue-green central part of the flame is about half the height of the whole flame. If the flame is very long and is yellow in parts, too little air is being admitted; if the flame is short and makes a slight roaring noise, too much air is being admitted. The amount of air admitted to a gas burner is usually adjusted by means of a valve or damper situated at the base of the burner. In all gas burners the various openings should be kept clean.

The heats of combustion of the fuels mentioned above are given in the following table. See also Table XVI, Appendix.

COMPARISON OF HEAT VALUES OF FUELS

	B.t.u. per lb.		*B.t.u. per lb.*
Wood, soft	7,000	Fuel oil	18,000
Wood, hard	7,000	Kerosene	19,980
Coal, bituminous	14,000	Gasoline	20,250
Coal, semi-bituminous	15,000		*B.t.u. per cu. ft.*
Coal, hard	13,000	Natural gas	1,000
Charcoal	16,000	Manufactured gas	600
Coke	12,000		

278. The Oxy-Acetylene Flame. Acetylene (C_2H_2) is a gas which results from the chemical reaction between calcium carbide (CaC_2) and water (H_2O). If crystals of calcium carbide are put into water, a clear, colorless gas (acetylene) is given off which burns with a brilliant white flame. Originally calcium carbide was expensive, but in 1892 it was discovered that it could be cheaply made by fusing lime and coal together in the intense heat

of an electric furnace. As a result of this discovery, calcium carbide was soon made on a large scale and sold at a moderate price.

The *oxy-acetylene torch* is a device by means of which oxygen and acetylene, contained in metal tanks under high pressure, is mixed and burned. The oxy-acetylene flame is exceedingly hot, and for this reason it is used

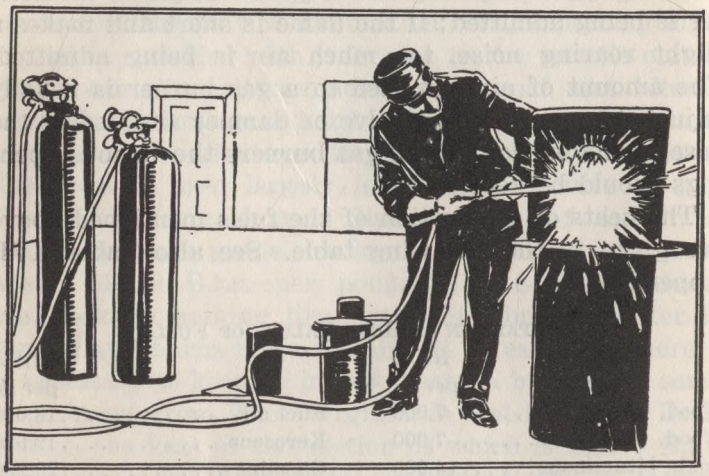


FIG. 233.—Cutting an iron plate with an oxy-acetylene flame

for cutting iron and steel bars and plate, Fig. 233, and for welding purposes.

279. Open Fireplace and Stove. Heating by means of an open fireplace has two things in its favor. First, the open fire gives to the room a cheery appearance, and second, it furnishes an excellent means of ventilation. A strong draft (convection current) up the chimney is created by the fire. This heated air is replaced by fresh air which finds its way in by doors and windows. Since the convection currents that are created are directed into the chimney, all the heat that gets out into the room must be

transmitted by radiation. The amount of heat sent out into the room by radiation is small as compared with the quantity which is carried up the chimney. It follows, therefore, that this method of heating is a very wasteful one.

A stove is a more effective heating device than the open fireplace, because heat is transmitted to the room by conduction and convection as well as by radiation. The stove, however, must be classified as a low grade heating system because of its relatively poor heat transmission and also because it fails to possess adequate ventilating facilities.

280. The Furnace. The ordinary hot air furnace is in principle nothing but a stove surrounded by a jacket and placed below the floor. The furnace has an advantage over the stove in that it may be more conveniently taken care of because of its location in the basement. And another advantage lies in the fact that pure out-door air may be supplied to the rooms above it. Furnaces are usually supplied with a water pan which should be kept full so as to keep the heated air from becoming too dry. In operating a furnace the following conditions should be noted: (a) It is good economy to keep the furnace well filled with fuel; that is, it is better to fill the furnace full once or twice a day than to put in smaller amounts

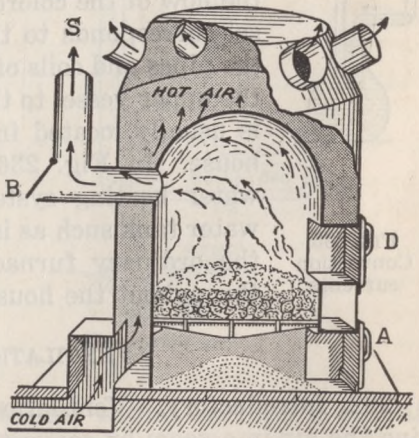


FIG. 234.—Hot air furnace

several times during the day. (b) In burning coal, the grates should be kept clean of ashes and clinkers. (c) If coke is used, a bed of ashes may remain on the grates below the fuel. Since coke burns with a very intense heat, care must be observed in regulating the drafts. (d) For a hot fire, open draft *A*, Fig. 234, and close the check draft *B;* for a slow fire, close *A* and open *B;* for medium fire, close both *A* and *B*. In the case of coke after the fire gets well started, close draft *A* and open *B*.

281. Use of Convection Currents in Heating. Convection currents such as are used in hot water heating systems may be demonstrated by a simple device, such as is shown in Fig. 235. Fill the flask and tubes with water. Then add to the open vessel at the top some colored liquid. Now when heat is applied to the lower vessel, the cold water in the tube descends and the hot water rises, thus giving rise to convection currents, as revealed by the flow of the colored liquid. The lower vessel corresponds to the furnace, the tubes to the pipes and coils of the heating system, and the upper vessel to the expansion tank which is usually located in the upper story of the house. In Fig. 236 there is shown a hot water heating system, and Fig. 237 a hot water tank such as is used in connection with the ordinary furnace to provide hot water throughout the house.

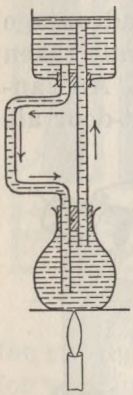


FIG. 235
Convection currents

VENTILATION

282. Introductory. Ventilation is the process of changing the air in a room by convection currents, as through open fireplaces, or ventilating flues, currents of air from open doors and windows, or by forced draft from fans.

To secure ventilation, provision must be made for an intake of fresh air and an output of foul air.

The fundamental principles governing ventilation may be illustrated by an experiment with a candle and a glass flask, Fig. 238. By means of a wire lower a small lighted candle into a flask having a rather wide neck. The flame burns brightly for a while, and then begins to grow dim, and soon goes out due to the

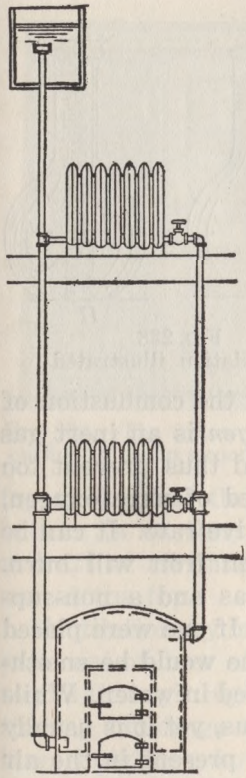


FIG. 236
Hot water heating system

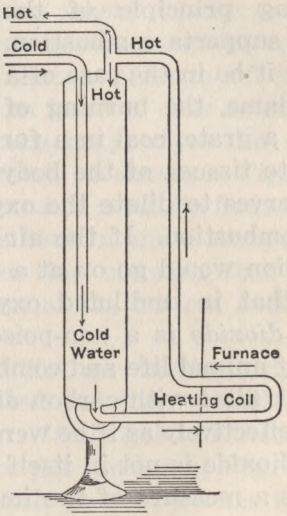


FIG. 237
Hot water tank

fact that the oxygen of the air within the flask is burned out. Now if a strip of tin or cardboard be thrust down into the neck of the flask and be so adjusted that a stream of cool air will descend on one side and a stream

of hot air ascend on the other, we shall have the conditions necessary for the ventilation of the flask, and the candle flame will burn brightly.

Air is composed of two gases, oxygen (O) and nitrogen (N) which occur in the approximate ratio of 1 to 4 by volume. In addition to oxygen and nitrogen the air also contains other substances, such as water vapor, ozone, carbon dioxide, smoke, dust, germs, etc. *Oxygen* is the life-giving principle of the air. It supports combustion, whether it be in the case of a candle flame, the burning of wood in a grate, coal in a furnace, or the combustion of the waste tissues of the body. *Nitrogen* is an inert gas which serves to dilute the oxygen and thus prevent too rapid combustion. If the air consisted of pure oxygen, combustion would go on at a destructive rate. It can be shown that in undiluted oxygen even iron will burn. *Carbon dioxide* is a non-poisonous gas and a non-supporter of animal life and combustion. If one were placed in a tank filled with carbon dioxide, he would be smothered as effectively as if he were drowned in water. While carbon dioxide is not in itself poisonous, yet it is usually taken as a measure of the impurities present in the air of a room due to respiration and other exhalations from the body. If there be a relatively large quantity of carbon dioxide present in a room, it is assumed that the air is impure.

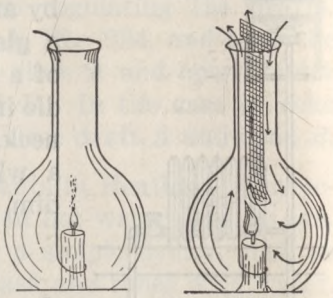


FIG. 238
Ventilation illustrated

The problem of the ventilation of rooms as affecting health and good working conditions is closely related to the problems of heating, air circulation, and humidity.

Proper air circulation and humidity conditions are of paramount importance.

283. Methods of Ventilating. Natural ventilation is that which takes place through the walls, cracks around doors and windows, and through the opening and closing of doors in passing from one room to another. Except on windy days this method is not sufficient to keep the air of most rooms in proper condition for breathing, and especially is this true of rooms where a large number of persons are congregated.

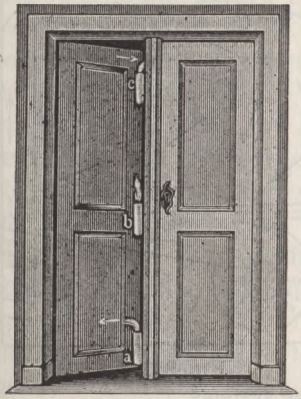


FIG. 239.—The inflow and outflow of air in a room

That a passage of air from room to room does actually take place may often be demonstrated, as shown in Fig. 239, by means of a candle flame. At the top of the door the flame is bent in one direction, at the bottom in another direction, thus showing that there is a draft of air into the room at the bottom and a draft out at the top.

Ventilation by means of hot air furnaces, in which pure out-of-door air is heated and then passed into the rooms through registers, is a fairly good method of ventilating buildings which are not too large, Fig. 240. In large buildings, however, such as auditoriums, the most economical and efficient method of changing the air is by means of mechanically driven fans, Fig. 241. Pure outdoor air passes over a system of steam heated pipes and is then driven by the fan into rooms of the building, entering near the ceiling and leaving at a point in the same wall below the inlet.

In buildings where relatively large numbers of persons

are assembled, it is estimated that the proper quantity of air, in cubic feet, per person per hour is about as fol-

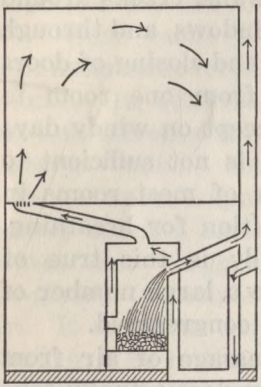


FIG. 240

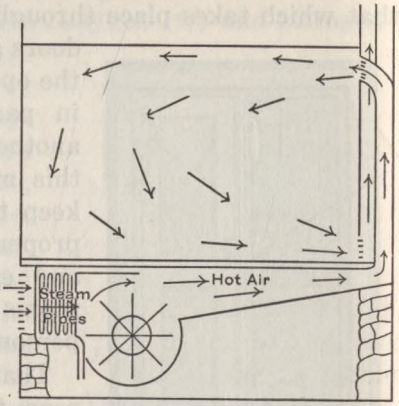


FIG. 241

Heating and ventilating systems

lows: (a) schools and churches, 2000; (b) theaters, 3000; (c) hospitals, 5000.

284. Relation of Humidity to Health and Bodily Comfort. Next to oxygen and nitrogen, water is one of the most important constituents of the air, and the role that it plays is of major importance in relation to bodily comfort and health. To maintain a healthful and comfortable condition of the body, it is necessary that air possess a relative humidity of about 60 per cent. The maintenance of about this degree of humidity is one of the objects of ventilation. If air is very dry, that is, if the humidity is low, there results an uncomfortable condition of the body due to the excessive extraction of water from the skin, and especially from the mucous membranes of the respiratory tract.

The proper degree of humidity not only prevents excessive moisture losses from the mucous membranes, but

it is also an important factor in the regulation of temperature of the body. The temperature of the normal healthy human body is very nearly constant at 98.5° F. Now one of the most important agencies in the regulation of bodily temperature is that of evaporation. Evaporation is a cooling process. When it goes on freely from the body, the temperature is lowered; when it is retarded, the temperature rises. When the air is filled with moisture, that is, when the humidity is high, evaporation is retarded, and hence the temperature of the body rises. Now when the air is dry, evaporation is accelerated and the body remains cool, even though the temperature is high. Thus it appears that our sensations of warmth and cold depend not only upon temperature but also on humidity as well. Low relative humidity then not only causes discomfort and lowers the tone and vigor of the body, but it also leads to a needless waste of fuel, variously estimated at from 10 to 25 per cent.

To conserve health, comfort, and fuel these conditions should be maintained: a relative humidity of about 60 per cent, and as far as possible a free circulation of the air.

To illustrate how the humidity of the air changes with a change of temperature, let us consider the following example: Suppose that out-door air in winter, having a temperature of 0° C and a relative humidity of 80 per cent, is passed through a furnace and is heated to 20° C (68° F), and thence passes to a living room. The phenomenon which is most apparent to the ordinary person is the change in temperature of the air from 0° to 20° C. This, however, is by no means the most important change which has taken place with reference to the air. For example, the air enters the furnace at 0° C, 80 per cent saturated, and emerges at 20° C. Referring to Table XIV, Appendix, we find that a cubic meter of

air at 0° C is capable of containing 4.8 grams of water vapor. A humidity of 80 per cent, therefore, will represent $4.8 \times 0.8 = 3.84$ grams of water vapor per cubic meter. Now the air is heated to 20° C, and its capacity for holding water is changed to 17 grams per cubic meter (Table XIV). Since relative humidity is the ratio of water vapor in the air at a given time to the quantity required for saturation, the relative humidity at 20° C is $3.84/17 = 22.6$ per cent. This means that the humidity of the air in passing through the furnace was lowered from 80 per cent to 22.6 per cent. For health and comfort, however, air should possess a relative humidity of about 60 per cent. It is evident, therefore, that unless some special provision be made for adding moisture to the air when thus heated, the humidity will be far too low.

Heat and Work

285. The Mechanical Equivalent of Heat. The first accurate determination of the relation between heat and work was made in 1845 by Joule, an English physicist.



Joule

Joule's experiment was conducted somewhat as follows: A given quantity of water was placed in a vessel containing a paddle wheel, Fig. 242. About the axle of the wheel there was wound a strong flexible cord to the end of which was attached a weight, W. When the weight was allowed to fall, the work done, measured in foot pounds, was equal to the weight W in pounds times the space s. The falling of the weight caused the paddles to rotate, thus doing work upon the water and causing a rise in temperature. Joule's experiment demonstrates that a weight of 77.2

pounds falling through a distance of 10 feet (representing work to the amount of $77.2 \times 10 = 772$ F. P.) would raise the temperature of 1 pound of water 1° F. This means that a pint of water (1 pound, approximately) in cooling from the boiling point (212° F) to room temperature (70° F) would give out heat enough to do more than 100,000 foot pounds of work. Thus: $(212 - 70) \times 772 = 109{,}624$ F.P.

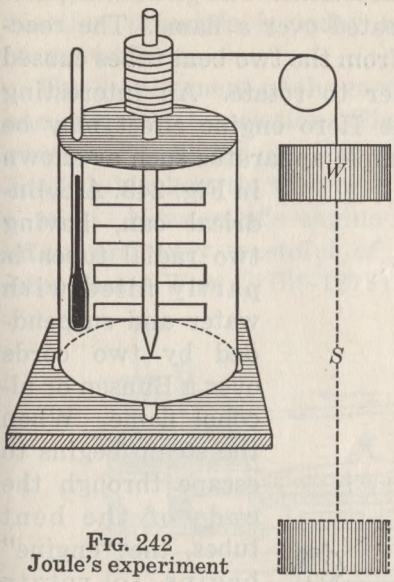


FIG. 242
Joule's experiment

The relation of work units to heat units is called the *mechanical equivalent of heat.* In 1879 Professor Rowland of Johns Hopkins University, Baltimore, made a series of classical experiments with an improved form of Joule's apparatus, by means of which he determined the following values for the mechanical equivalent of heat:

$$1 \text{ } calorie = 4.19 \times 10^7 \text{ } ergs$$
$$1 \text{ } B.t.u. = 778 \text{ } foot\ pounds$$

286. The Steam Engine. Joule transferred mechanical energy into heat energy. The heat engine is a device for accomplishing the reverse operation, namely, the transformation of heat into mechanical energy. The first heat engine was probably invented as early as 150 B. C. by Hero, a pupil of Archimedes. Hero's engine consisted of a hollow ball pivoted so as to rotate about a given axis.

Extending radially from the vessel were two tubes, their openings bent in opposite directions. The globe was partly filled with water and heated over a flame. The reaction of the steam escaping from the two bent tubes caused the hollow sphere or boiler to rotate. An interesting experiment illustrating the Hero engine effect may be performed by the use of simple apparatus such as shown in Fig. 243. A cylindrical can, having two radial tubes is partly filled with water and suspended by two cords over a Bunsen or alcohol flame. When the steam begins to escape through the ends of the bent tubes, the "engine" begins to rotate about its vertical axis. This rotation twists the suspending cords, and as a result the boiler is steadily lifted from the flame. When it has reached a certain elevation, steam will no longer be generated rapidly and the motion will cease. The boiler will now descend under the action of gravity, the cords in untwisting causing it to revolve in the reverse sense. Descending in this manner toward the flame, the boiler will eventually reach a position in which steam will again be generated, when it will once more climb out of reach of the flame. The reaction of the steam jets from the

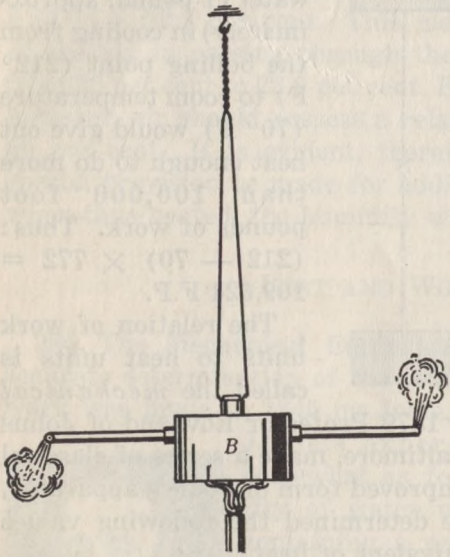


FIG. 243.—Demonstration of Hero's steam engine

nozzles of the tubes supplies the motive force which sets the system in rotation. It is interesting to note that this reaction principle is today used in some types of the steam turbine.

The development of the modern steam engine has been one of gradual evolution. The first successful steam engine was invented in 1705 by Thomas Newcomen, an English blacksmith, who used it in pumping water from a mine. Newcomen's engine was mechanically a crude affair and very wasteful of fuel. Some seventy years later James Watt (1736-1819), a Scotch instrument mak-

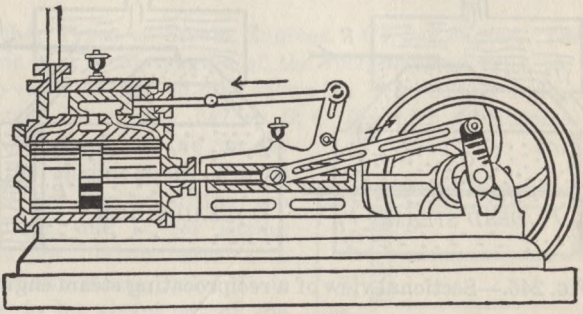


FIG. 244.—Reciprocating steam engine

er, invented an improved type of engine that was so superior to any of the earlier types that he is often given the credit of being the inventor of the steam engine. For the same amount of work done, Watt's engine used only about one fourth as much coal as its predecessor.

287. The Reciprocating Steam Engine. Steam engines are of various types. With reference to the nature of the principal driving parts, steam engines may be classified as (a) reciprocating, and (b) rotary. Again, as to the disposal of the exhaust steam, that is, the steam which has been used in driving the moving parts, steam engines may be classified as (c) noncondensing, and (d) condensing engines. And, finally, with reference to the num-

ber of expansions performed by the steam, engines are (e) simple and (f) compound.

The principle by which expanding steam is made to do work can very well be illustrated in an elementary way by means of a study of the simple, noncondensing, reciprocating engine. A *reciprocating engine* is one in which the piston or driving part moves back and forth in a cylinder. The working parts of such an engine are shown in Fig. 244 and by means of line drawings in

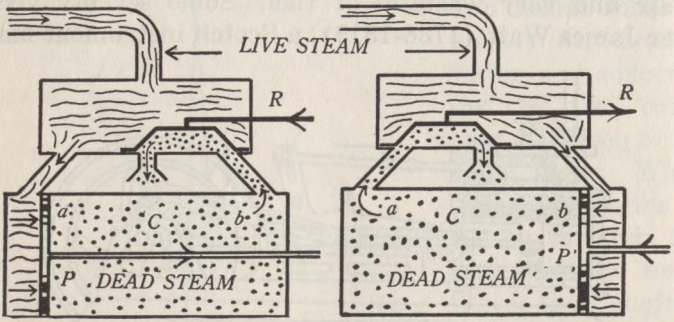


FIG. 245.—Sectional view of a reciprocating steam engine

Fig. 245. Steam under high pressure is admitted to the cylinder *C*, first on one side and then on the other of the piston *P*, causing it to move back and forth from *a* to *b*. The admission of the steam to the cylinder is regulated by means of the slide-valve *V*, which is attached to the eccentric by means of the rod *R*. An *eccentric*, Fig. 246, is a circular disc which is rigidly attached to the fly wheel shaft but so set that its center does not coincide with the center of the shaft. When the fly wheel shaft rotates, the rotation of the eccentric communicates a smooth, back

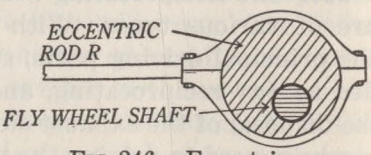


FIG. 246.—Eccentric

and forth motion to the eccentric rod *R*. The motions of the sliding valve *V* and that of the piston *P* are always in opposite directions. In *I*, Fig. 245, the piston is represented as moving to the right; in *II*, to the left.

The steam engine, with its boiler, is a machine for transforming the potential energy of coal into the kinetic energy of mechanical motion. The two essential parts of the complete mechanism are the boiler, in which the energy of the coal is transformed into the energy of steam under pressure, and the engine in which the energy of the steam is transformed into mechanical motion.

288. Other Types of Steam Engines. *Corliss engine.* The Corliss engine is a steam engine of the reciprocating type, in which the slide-valve of the engine shown in Fig. 244 is replaced by four rotating valves, Fig. 247, *A*, *B*, *C*, *D*, which are automatically opened and closed at the proper time by the engine itself. Valves *A* and *B* are for admitting the steam, and *C* and *D*, for letting the steam out. When valve *A* is open to admit steam, valve *D* is also open to let steam out of the other end of the cylinder, while *B* and *C* are closed; on the reverse stroke *B* and *C* are open, while *A* and *D* are closed. Corliss engines are more efficient than reciprocating engines of the slide valve type, and are used extensively in large plants.

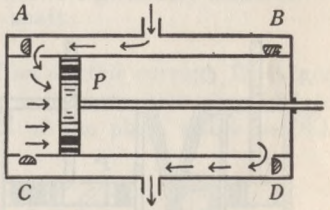


FIG. 247.—Corliss engine

Condensing engine. Most of the heat brought to the engine by the steam from the boiler is ejected by the engine in the exhaust. This loss of heat varies from 70 per cent of the heat of the steam in the best engines, to over 90 per cent in the poorer types. In many manufacturing establishments the heat of this exhaust steam is recovered by using it for heating or manufacturing purposes. When a vapor, not in contact with its own liquid, is heated so that it possesses more than enough heat to maintain its existence as a dry gas at a given pressure, it is said to be superheated. In some modern engines steam is superheated to the condition of a

dry gas previous to its admission to the engine. This is to prevent its condensation on entering the expansion chamber. If the steam after doing work on the piston is allowed to escape directly into the air, it is clear that the piston will always be working against the back pressure of the atmosphere. That is to say, there will be two forces acting on the piston when it is in motion; one is the steam tending to drive it in one direction, and the other is atmospheric pressure tending to drive it in the opposite direction. It is clear, therefore, that if the atmospheric pressure could be entirely removed, the efficiency would be increased to the extent of 14.7 pounds per square inch. The *condensing engine* is one in which the exhaust steam passes into a cooling chamber. This usually consists of a chamber surrounded by water and into which a spray of cold water is admitted. When the steam comes in contact with the cold spray it is condensed, producing a partial vacuum, thus reducing the back pressure on the piston.

Compound engine. A compound engine is one in which the steam under high pressure enters one cylinder, does work upon the piston, then enters a second cylinder, again doing work, Fig. 248. It will be noted that there are six rotating valves, *A*, *B*, *C*, *D*, *E*, and *F*, connected with the two cylinders, *A* being the high pressure cylinder and *B* the low pressure. The low pressure cylinder is considerably the larger of the two.

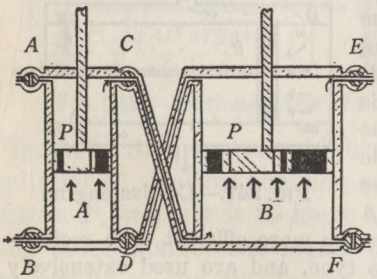


FIG. 248.—Compound engine

When the steam exerts its pressure upon two cylinders in succession, the engine is called a double-expansion engine; when it acts upon three cylinders, a triple-expansion engine, etc. In some of our ocean going vessels, engines have been used of the triple-expansion type.

289. The Steam Turbine. The steam turbine, Fig. 249, works on the same principle as that of the wind mill or the water turbine, except that the energy is furnished by steam instead of wind or water. Steam under high pressure is directed by one or more nozzles against a series of curved blades set around the outer margin of a wheel, Fig. 250. Since this wheel rotates at an enor-

mously high speed, 50,000 to 70,000 r.p.m., the shaft cannot be connected directly to the driving shaft of the ordinary machine. Contact with the shaft of the turbine is made through reduction gears or by means of an electric generator, which transforms the

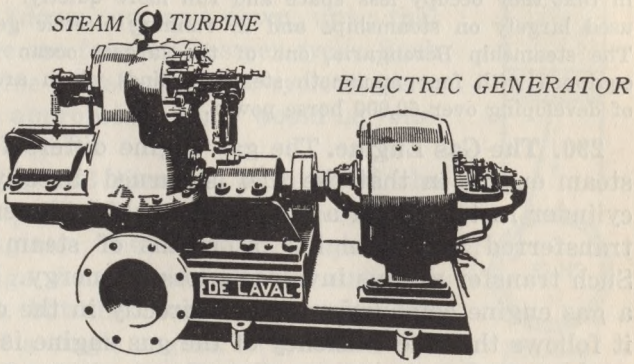


FIG. 249.—Steam turbine with an electric generator attached to the drive shaft

energy of rotation into the energy of an electric current. By means of electric motors this electric energy is again transformed into mechanical energy in different parts of the plant where needed.

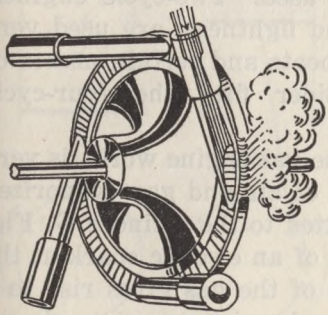


FIG. 250

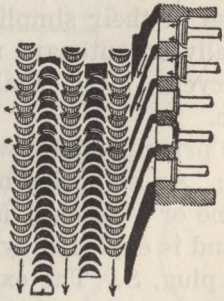


FIG. 251

Sectional view of steam turbine

Instead of directing the steam against one set of blades or buckets, it is usually passed through a series of curved blades which alternate with a corresponding series of stationary curved blades, as

shown in Fig. 251. The black vanes are fastened rigidly to the steam chamber, called the stator; the light ones are attached to the shaft, and constitute the rotor.

Steam turbines have an advantage over reciprocating engines in that they occupy less space and run more quietly. They are used largely on steamships and in running electric generators. The steamship Berengaria, one of the largest ocean liners, is equipped with four mammoth steam turbines which are capable of developing over 60,000 horse power.

290. The Gas Engine. The gas engine differs from the steam engine in that the fuel is burned directly in the cylinder instead of in a boiler from which the energy is transferred to the engine by means of steam pipes. Such transfer always involves a loss of energy. Since in a gas engine combustion occurs directly in the cylinder, it follows that the efficiency of the gas engine is greater than that of the steam engine. Gas engines are known technically as *internal combustion engines.*

Gas engines are of two general types, the *four-cycle* engine and the *two-cycle* engine. Where high power and high efficiency are required, as in automobiles, for example, the four-cycle engine is used. Two-cycle engines, because of their simplicity and lightness, are used very generally on outboard motor boats and certain makes of motorcycles. We shall consider first the four-cycle engine.

The principle upon which the gas engine works is very simple. An explosive mixture of air and gas (vaporized gasoline or kerosene) is admitted to the cylinder *G*, Fig. 252, and is exploded by means of an electric spark at the spark plug, *S*. The explosion of the gas gives rise to a sudden expansion which causes the piston to move downward. The inertia of the flywheel carries the piston back to the upper end of the cylinder, expelling the waste gas through the valve *V'*. On the next half revolution of the flywheel the piston moves down, sucking in

a new charge of gas through *V*. As the flywheel completes its revolution, this gas is compressed and exploded. Thus from one explosion to the next (one cycle) there are four single strokes of the piston (down, up, and down, up); that is, there are four strokes per cycle. Such an engine is called a four-cycle engine; a more appropriate name would be *four-stroke* engine.

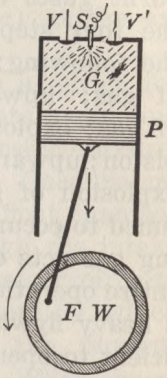


FIG. 252
Section of four-cycle gas engine

In Fig. 253 there is represented the four steps of a four-cycle, single cylinder gas engine. The student should note that Figs. *I* to *IV* do not represent four engines, but the four steps during one cycle of a single engine. (a) First, the flywheel is turned by hand, or other device, causing the piston to move down and thus sucking in a mixture of air and gas through the

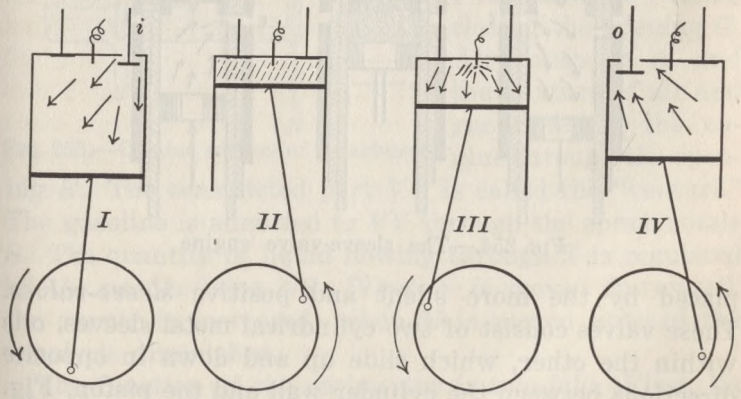


FIG. 253.—Illustrating the four steps of a four-cycle, single-cylinder gas engine

valve *i*. (b) As the flywheel continues its rotation, the piston moves back, compressing the gas. (c) The

compressed gas now explodes, driving the piston forward. (d) On the back stroke of the piston the burnt gases are ejected through valve *o*. To repeat, the four steps required to complete the cycle include the following: Downward stroke of piston and intake of gas; upward stroke of piston and compression of gas; explosion of gas, causing downward motion of piston; upward motion of piston and ejection of gas. The explosion of gas is due to an electric spark, which is timed to occur at the proper moment. Since the expanding gas acts on the piston during only one part of this entire operation, it is necessary in gas engines to provide a heavy flywheel, the stored up energy of which is sufficient to operate the piston during the remaining three strokes.

In the Knight engine, "poppet" valves have been re-

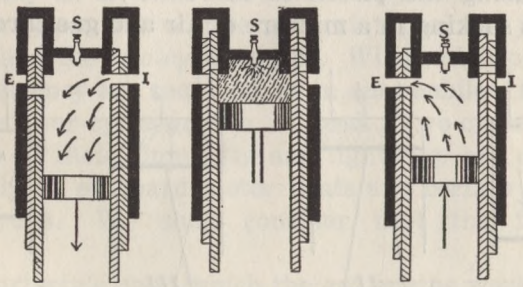


FIG. 254.—The sleeve-valve engine

placed by the more silent and positive *sleeve-valves*. These valves consist of two cylindrical metal sleeves, one within the other, which slide up and down in opposite directions between the cylinder wall and the piston, Fig. 254. Each valve has two openings which, when they are in alignment, allow gas to be drawn into the cylinder at the intake port *I* and the burned gas to be expelled through the exhaust port *E* once during each cycle. The

operation of the valves is as follows: (a) At the intake stroke of the piston, gas is drawn into the cylinder, Fig. 254, left; (b) during the compression and explosion strokes, both ports are closed, middle; (c) at the exhaust stroke, port *E* is open, right.

291. The Carburetor. The *carburetor* is a device for converting liquid fuel, such as gasoline, into vapor and mixing it with the air to make an explosive mixture suitable for complete combustion, Fig. 255. Gasoline enters the float chamber *R*, through the intake *G*. This reservoir contains a float *FF* which controls the plunger valve *P*. When chamber *R* is full, *FF* rises, operating the levers *LL* and thus causing *P* to move downward closing the opening *G*. Air enters at *A*, and the mixture of air and gas flows to the engine through the opening *E*. The constricted part *VV* is called the "venturi." The gasoline is admitted to *VV* through the spray nozzle *S*. The quantity of liquid flowing through *S* is regulated by the needle valve *NS*. When *N* is moved downward, the supply is increased; when *N* is moved upward the supply is diminished.

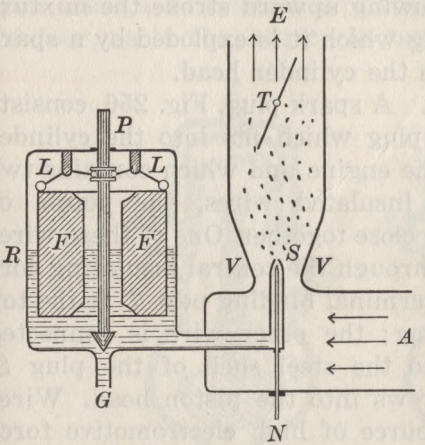


FIG. 255.—Outline section of a carburetor

The function of the carburetor is two-fold. First, its function is to supply a fine spray of gasoline into the venturi *VV*; second, its function is to regulate the supply of gasoline so that the ratio by weight of gasoline to air is about 1 part of gasoline to 15 parts of air, this

being approximately the right proportion for perfect combustion. Naturally the question arises here, How does air get into the carburetor, and how does the gas-air mixture get into the engine? When the piston head, Fig. 253, *I*, moves down, a partial vacuum is created in the cylinder. If now the throttle *T*, Fig. 255, is open, air is sucked into *A* and a spray of gasoline flows from *S*, forming an explosive mixture in *VV*. This mixture is drawn into the engine by the initial downward stroke of the piston. On the following upward stroke the mixture is compressed, following which it is exploded by a spark from the spark plug in the cylinder head.

292. The Spark Plug. A spark plug, Fig. 256, consists essentially of a metal plug which fits into the cylinder head of the engine and which contains two carefully insulated wires, the points of which are close together. One of these wires runs up through the central insulating core *C* to the terminal binding post *T* at the top of the plug; the other wire is connected directly to the steel shell of the plug *S*, which screws into the piston head. Wires from a source of high electromotive force (battery or magneto and induction coil) are connected to the plug as follows: One wire from the battery system is connected directly to the terminal binding post *T*, and the other wire is connected to the metal case of the engine, the current being conducted to the plug wire through the metal shell *S*. It thus appears that while only one wire from the battery is visibly connected to the spark plug at the terminal *T*, the other connection from the battery is made through the metal parts of the engine.

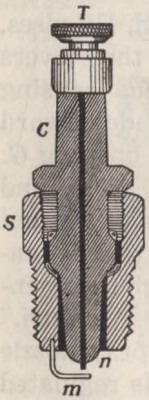


FIG. 256
Section of a spark plug

When the electric circuit is closed through the mechan-

ism of the ignition system of the engine, a spark occurs at the spark tips *mn*, thus exploding the mixture of air and gasoline in the cylinder. The essential parts of the ignition system of a gas engine consist of a battery or magneto (Art. 436), an induction coil (Art. 422), and a rotating device which is driven by the engine, and which is so designed as to open and close the electric circuit through the spark plug at the proper intervals.

293. The Two-Cycle Engine. In the *two-cycle* gas engine there are two strokes of the piston per cycle instead of four, as in the four-cycle machine, that is to say, in

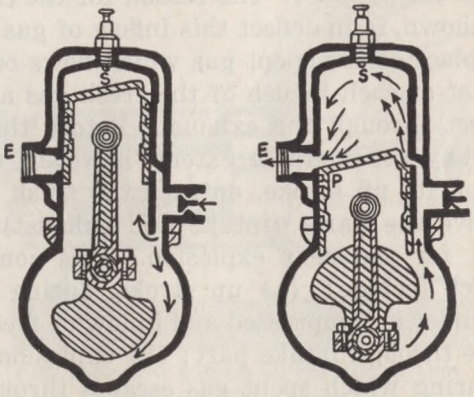


FIG. 257.—Sectional view of a two-cycle gas engine

the two-cycle engine there is one explosion of the gas for every revolution of the flywheel. A study of Fig. 257 will reveal three special features about the design of this engine: first, there is only one valve, and that at the intake, second, the crank case below the piston rod is air tight, being a continuation of the supporting part of the cylinder case and water jacket, and third, the upper face of the piston slants downward from the intake side toward the exhaust side. The reasons for this special de-

sign will appear as we study the operations involved in a single cycle.

Start with the piston at the end of its up stroke, Fig. 257, left, with the explosive mixture of gas and air compressed above it. During this up stroke a charge of gas is drawn into the crank case through the intake. Now when the gas above the piston explodes, the piston moves down, Fig. 257, right. When the piston head passes below the opening of the exhaust pipe E, the spent gas escapes. At the same instant, however, the intake valve closes and the slightly compressed gas in the crank case rushes into the cylinder. The reason for the tilted piston head, as shown, is to deflect this inflow of gas upward to take the place of the spent gas which flows out at E. It is true that at each inrush of the fresh gas a little of it will be lost through the exhaust. Before this loss can occur to any considerable extent, however, the piston, starting on its up stroke, entraps the fresh gas in the space above the parts (intake and exhaust) and compresses it for the next explosion. This completes the cycle which include: (a) up stroke, during which gas above cylinder is compressed and inflow of fresh gas into crank case through intake part; (b) explosion and down stroke, during which spent gas escapes through the exhaust and gas from crank case flows into cylinder.

The main disadvantage of the two-cycle engine is that it requires more fuel per horse power than the four-cycle type because some gas is lost with each exhaust. Its advantages lie in the fact that it is very light per horse power, and that it delivers a positive thrust for each revolution of the flywheel.

294. The Oil Engine. This is an internal combustion engine in which oil of a more or less crude nature is the fuel used. The most efficient oil engines are of the Diesel or semi-Diesel types. The principle of operation is as

follows: Air is drawn into the cylinder on the intake stroke and on the succeeding forward stroke is compressed to a pressure of 450 or 500 pounds per square inch. This compression raises the temperature of the air to the firing point. At the end of the compression stroke oil, under a pressure still greater than that in the engine, is sprayed into the cylinder. This mixture of oil vapor and air is ignited by the high temperature produced during the preceding compression stroke. The resulting expand-

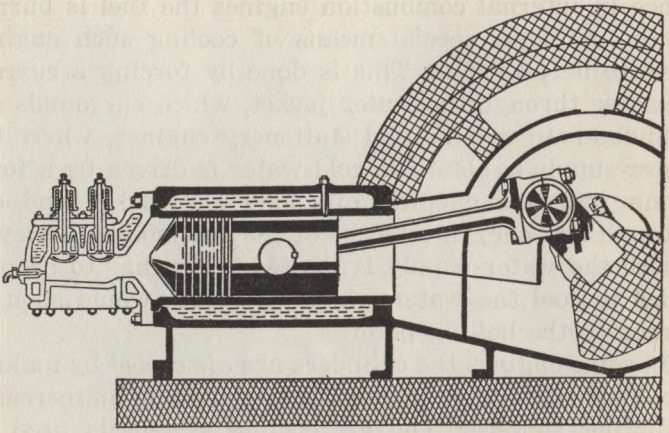


Fig. 258.—Oil engine of the Diesel type

ing gas pushes back the piston as in a manner similar to that of the steam engine. In the Diesel engine the oil burns non-explosively, maintaining during the first part of the stroke a pressure which is of practically constant value; in the semi-Diesel type, Fig. 258, the oil burns much more rapidly, the combustion being almost explosive in nature.

The characteristic features of oil engines of the Diesel type are as follows: (a) Injection of fuel at the end of the compression stroke; (b) self ignition of the charge by the high temperature of the compressed air; (c) no

carburetor or spark plugs; (d) the highest efficiency of any heat engine thus far devised.

Engines of the Diesel type are used in submarines because of their economy of fuel and because of absence of smoke and the consequent elimination of large funnels. They are also rapidly coming into general use on steam ships and for many land operations where high power service is required.

295. Cooling Systems for Internal Combustion Engines. Since in internal combustion engines the fuel is burned in the cylinder, special means of cooling such engines have to be provided. This is done by forcing a current of water through the water jacket, which surrounds the cylinders. In marine and stationery engines, where the water supply is plentiful, cold water is driven by a force pump into one opening of the water jacket and out through another; in the case of the automobile, however, where the water supply is limited, a fan has to be provided to cool the water which otherwise would soon be heated to the boiling point.

In some engines the cylinders are *air cooled* by making the outer surface corrugated, thus providing an increased radiating surface. The air cooling system is used on motorcycle and airplane engines; it is also a feature of certain automobiles.

296. Efficiency of Heat Engines. The efficiency of any piece of heat apparatus, whether it be a steam engine, a furnace, or an ordinary gas burner under a kettle, is the ratio of the useful energy gotten out to the total energy put in; that is,

$$efficiency = \frac{useful\ energy\ out}{heat\ energy\ in} = \frac{output}{input}$$

Example. One cubic foot of gas is burned under a kettle containing a gallon (8 pounds) of water. The temperature of the water is changed from 68° F to 110° F.

Find the efficiency of the burner and kettle. *Solution*: One cubic foot of gas gives 600 B.t.u. of heat = input. To change 8 pounds of water from 68° F to 110° F (110 — 68 = 42) requires 8 × 42 = 336 B.t.u. = output. Therefore the efficiency = 336/600 = 0.56 = 56 per cent.

As applied to heat engines, the term *efficiency* may be used in several different senses, as for example boiler efficiency, engine efficiency, the thermal or fuel-engine



FIG. 259.—In the ordinary railway locomotive only about 7 per cent of the energy in the coal is converted into useful work

efficiency. (a) *Boiler efficiency* is the ratio of the heat energy in the steam generated in the boiler to the energy contained in the fuel used. In general, boiler efficiency is from 60 to 80 per cent. (b) *Engine efficiency* is the ratio of the mechanical work done on the flywheel shaft to the heat energy delivered by the steam. (c) *Fuel-engine* efficiency includes the efficiency reckoned from

fuel to drive shaft. In the case of the steam engine, it takes account of both boiler and engine efficiency. For example, suppose that the boiler efficiency of a certain boiler-engine system is 60 per cent, and that the engine efficiency is 18 per cent, then the fuel-engine efficiency is $.18 \times .60 = 0.108 = 10.8$ per cent. (d) Efficiency is also sometimes expressed in pounds of coal required for each horse power hour.

The fuel-engine efficiency of the railway locomotive is relatively very low. The average efficiency of the ordinary locomotive is from 7 to 8 per cent, which means that for every 100 tons of coal used the energy of 90 or more tons is wasted. The locomotive is an exceedingly wasteful device, so far as fuel energy is concerned. A glance at Fig. 259, in which there are shown five engines, four pulling and one pushing, a train up a heavy western mountain grade, will explain in part why the efficiency of the ordinary locomotive is low. Not only is energy wasted from smoke stack and safety valve, but by heat radiation from the boiler as well.

297. The Ideal Engine. An *ideal engine* is one in which all the processes may be conceived of as being carried on without loss of any sort, as by radiation, conduction, or any other means whatsoever. No such engine actually exists because, of course, there is always bound to be some heat loss. This ideal engine is called a *Carnot's engine*, from the name of the man (Sadi Carnot, a French physicist and engineer) who first used the idea as the basis for the development of certain important mathematical equations expressing the relationship between heat and work. He showed that the efficiency of an ideal engine is

$$E = \frac{T - T'}{T}$$

in which T is the temperature in *absolute* units of the medium (steam, say) received by the engine and T' is the absolute temperature of the medium discharged. No engine can have an efficiency greater than this fraction $(T - T')/T$.

Example. By way of illustration let us suppose that a steam

engine receives superheated steam from the boiler at a temperature of 200° C and delivers the spent steam to a condenser at 40° C. Find the efficiency of an ideal engine working between these temperature limits. *Solution*: 200° C = 473° *Absolute*, and 40° C = 313° *Absolute*. Then for these absolute temperature limits

$$\textit{efficiency} = \frac{473 - 313}{473} = 0.338 = 33.8 \textit{ per cent}$$

In actual practice, then, no engine, *working between the temperature limits given*, can have an efficiency greater than this; in fact, the actual efficiency for these temperature limits (200° C and 40° C) is only about 25 per cent. The student should bear in mind that in this case we are speaking of *engine efficiency*, no account being taken of the boiler losses. An engine that had an engine efficiency of 25 per cent and a boiler efficiency of 70 per cent would have a *fuel-engine* efficiency of .25 × .70 = 0.175 = 17.5 per cent.

The fuel-engine efficiency of gas and oil engines is relatively high, because the fuel is burned directly in the engine, boiler and transmission losses being thus eliminated.

In the following table there are given the approximate thermal (fuel-engine) efficiencies of different types of heat engines.

THERMAL EFFICIENCIES

Non-condensing engine	9%	Steam turbine	26%
Condensing engine	17%	Gas engine	28%
Quadruple expansion engine	24%	Oil engine, Diesel	36%

SUMMARY

1. Cooling effects may be produced by (a) melting of ice, (b) solution of salts in liquids, (c) evaporation of liquids, and (d) expansion of gases.

2. Each gram of ice melted takes up 80 calories of heat; each pound, 144 B.t.u.

3. Heat of combustion of a fuel is the number of heat units liberated (calories or B.t.u.) per unit mass (gram or pound) of the fuel consumed.

4. Comfortable and healthful living conditions in rooms depend on temperature, ventilation, and relative humidity. The relative humidity in living rooms should be about 60 per cent.

5. Mechanical equivalent of heat: 1 calorie = 4.19×10^7 ergs; 1 B.t.u. = 778 foot pounds.

6. Steam engines may be classified according to the working parts as reciprocating and rotary; according to exhaust, as non-condensing and condensing; according to steam expansion, as simple or compound.

7. Internal combustion engines are classified according to fuel uses as gas engines and oil engines. Gas engines may be four-cycle or two-cycle.

8. The term cycle, as applying to internal combustion engines, includes the period from one explosion to the next.

9. In the four-cycle engine there are four strokes of the piston per explosion, two up and two down; that is, the flywheel makes two revolutions per explosion.

10. In the two-cycle engine there are two strokes of the piston, one up and one down; the flywheel makes one revolution for each explosion.

11. The efficiency of an engine (not including the boiler) is the ratio of the mechanical energy delivered to the available energy supplied.

12. Fuel-engine efficiency is the ratio of the mechanical energy given out to the heat energy generated by the burning of the fuel. In the case of the steam engine, this includes both the boiler efficiency and the engine efficiency.

13. An ideal engine is an imaginary engine in which there are no losses by conduction or radiation. The efficiency of such a theoretical engine is $E = (T - T')/T$, in which T is the absolute temperature of the working substance (steam or gas) and T' is the absolute temperature of the exhaust.

14. The efficiency of an engine depends only on the difference between the temperatures at intake and exhaust; that is, upon $T - T'$.

15. For a given temperature range between intake and exhaust, no actual engine can have an efficiency greater than that expressed by the equation $E = (T - T')/T$.

A

1. Give three methods of producing cold by artificial means, and give an example illustrating each.

2. Explain the ice cream freezer with reference to construction and refrigeration.

3. Name some of the fuels used in your community, and give their respective heats of combustion.

4. Three tons of hard coal of average grade are burned. (a) How many B.t.u. of heat are liberated? (b) How many cu. ft. of illuminating gas would be required to furnish the same quantity of heat?

5. Suppose that 10 tons of semi-bituminous coal are required to heat a certain house for a year, the heating apparatus being a hot air furnace, the efficiency of which is 60 per cent. Find (a) the total number of B.t.u. liberated in the furnace; (b) the number of heat units delivered through the registers.

6. What are some of the advantages of a furnace (hot air, hot water, or steam system) over those of a stove for heating a residence?

7. Discuss the care and operation of a heating furnace, with reference to economy in the use of different fuels.

8. Explain why buildings used for animal habitation (human or otherwise) should be ventilated.

9. Discuss several different methods of ventilation.

10. Why is a low relative humidity in living rooms conducive to discomfort and ill health?

11. Explain why a low relative humidity of the air in a room makes it seem colder than it really is.

12. Why is it desirable that the air in a living room should be kept in circulation?

13. Describe Joule's experiment.

14. Define mechanical equivalent of heat, and give its value in (a) metric units; (b) English units.

15. One quart of water (2 lbs.) is heated from the freezing point to 132° F. (a) How many B.t.u. are required to effect this change of temperature? (b) The heat consumed by the water is equivalent to how many foot pounds?

16. By means of a sketch explain the operation of a slide valve reciprocating engine.

17. Explain the meaning of the term cycle, as applying to internal combustion engines.

18. What is the distinction between a four-cycle and a two-cycle engine?

B

19. The temperature of 10 grams of water is changed from 20° C to the boiling point. Find (a) the number of calories of heat

required; (b) the equivalent value in ergs; (c) the equivalent value in joules?

20. By means of a sketch explain what takes place in the cylinder of a four-stroke gas engine during one cycle.

21. By means of a sketch explain the function of a carburetor.

22. What is meant by (a) boiler efficiency? (b) engine efficiency? (c) fuel-to-fly-wheel efficiency?

23. In one hour a 10-horse power engine burns 18 lb. of bituminous coal. (a) How much work in foot pounds is done by the engine during the hour? (b) The burning of 18 lb. of coal is equivalent to how many B.t.u.? (c) 18 lb. of coal is equivalent to how much energy in foot pounds? (d) What is the efficiency of the engine?

C

24. Explain the meaning of each term in the Carnot equation $E = (T - T')/T$.

25. The explosion temperature in a certain internal combustion engine is 1200° C and the temperature of exhaust is 700° C. What is the maximum theoretical efficiency of this engine?

26. Explain why the actual efficiency, such as found in Ex. 25, can never be as great as the theoretical efficiency.

NOTE: For additional problems, see Appendix.

CHAPTER IX

MAGNETISM

Magnets and Their Properties

298. Introductory. One of the oldest and at the same time one of the most interesting pieces of electrical apparatus is the magnet. The small horseshoe magnet with its mysterious power of attracting objects made of iron and steel, Fig. 260, is familiar to every school boy; and the study of the applications of magnetic attraction and

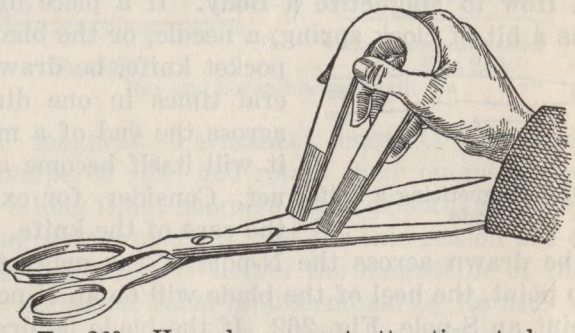


Fig. 260.—Horseshoe magnet attracting steel

repulsion, entering as these properties do into nearly every household and commercial electrical device, is of great interest and importance to the physicist and electrical engineer. The magnet was known and its properties applied by the Chinese at a very early date in the history of civilization; indeed, the magnetic compass was invented in China and was introduced into Europe from that country as early as 1300 A. D.

In many parts of the earth there is found a kind of

iron ore which possesses the properties of a magnet; that is, it attracts iron, and when suspended so as to swing freely about a vertical axis, it points in a north-south direction. This ore is called *magnetite* and its chemical composition is represented by the formula Fe_3O_4; that is to say, a molecule of magnetite contains 3 atoms of iron and 4 atoms of oxygen. A piece of this ore is called a *natural magnet* or *lodestone*. The term lodestone (leading stone) probably came from the fact that elongated pieces of magnetite were employed in early times by navigators as a sort of crude mariner's compass, Fig. 261.

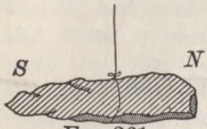


FIG. 261
Lodestone pointing in a north-south direction

299. How to Magnetize a Body. If a piece of steel, such as a bit of clock spring, a needle, or the blade of a pocket knife, be drawn several times in one direction across the end of a magnet, it will itself become a magnet. Consider, for example, the case of the knife. If the blade be drawn across the N-pole of the magnet from heel to point, the heel of the blade will be an N-pole and the point an S-pole, Fig. 262. If the blade be drawn in a similar manner across the S-pole of the magnet, the polarity of the blade will be reversed. The greater the number of times the steel is drawn across the magnet, the greater will be the pole strength, up to a certain point at which the steel is said to become magnetically saturated.

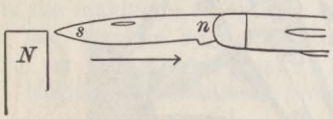


FIG. 262.—Magnetizing a knife blade

300. Magnetic Substances. If the pole of a magnet be brought successively in contact with pieces of iron, copper, brass, zinc, lead, silver, gold, graphite, wood, and paper, it will be found that the iron is strongly attracted,

while the other substances are not noticeably affected. The iron is magnetic; the others are non-magnetic.

A *magnetic substance* is one that is attracted or repelled by a magnet. Iron is highly magnetic; cobalt and nickel are also magnetic, but to a less degree than iron. Some magnetic substances, antimony (Sb) and bismuth (Bi) for instance, are slightly repelled by a magnet and are said to be *diamagnetic.*

301. Kinds of Magnets. Magnets may be classified in various ways, as for example: (a) natural and artificial magnets; (b) permanent and temporary magnets; and (c) with reference to their form, bar and horseshoe magnets, Figs. 263 and 264. All manufactured magnets are

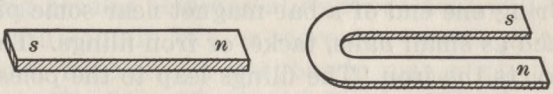


FIG. 263 FIG. 264

Bar and horseshoe magnets

artificial magnets. Permanent magnets are made of highly tempered steel and retain their magnetic properties for a long time; magnets made of soft iron soon lose their magnetic properties, and for this reason are called temporary magnets. Whether the magnet be of the bar form or horseshoe form is determined by its use. When

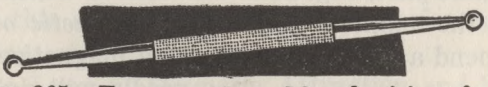


FIG. 265.—Eye magnet, used by physicians for removing particles of iron or steel from the eye

we desire to use a single pole, the bar magnet is the more convenient; when, on the other hand, the maximum lifting effect is desired, the horseshoe type is employed. Sometimes a short piece of soft iron, called an armature or keeper, is placed across the ends of the magnet when

the latter is not in use. The reason for using a keeper is to prevent the magnet from losing its strength, as will be explained later.

In Fig. 265 there is shown an eye magnet, such as is frequently employed by physicians to remove particles of iron or steel from the eye.

The electric door bell, the telephone, the telegraph, and the electric generator and motor are only a few of the many modern appliances in which magnets are used in one form or another.

302. Characteristic Properties of the Magnet. The characteristic properties of the magnet may be demonstrated experimentally as follows: (a) *Magnetic attraction.* Bring one end of a bar magnet near some pieces of iron, such as small nails, tacks, or iron filings. The magnet attracts the iron. The filings leap to the poles of the magnet and cling tenaciously. (b) *Magnetic polarity.*

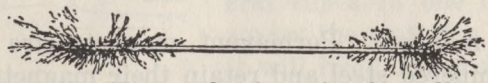


FIG. 266.—Magnet, showing tufts of iron filings clinging to the poles

Roll the magnet in iron filings; tufts of the filings cling to the ends of the bar, while few or none are attracted to the middle part, Fig. 266. The ends to which the filings cling are the magnetic poles. (c) *Magnetic orientation.* Now suspend a light bar magnet (a magnetized knitting needle will do) so as to move freely about an axis, Fig. 267. The magnetic needle takes a north-south position, the end pointing toward the north being called the *north-seeking pole*; the end pointing to the south, the

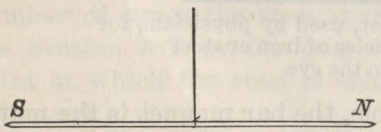


FIG. 267.—Magnetic needle, showing north- and south-seeking poles

south-seeking pole. The north-seeking pole is usually marked N or $+$; the south-seeking pole, S or $-$. A straight line joining the N and S poles of a magnet is called the *magnetic axis.*

303. Consequent Poles. Every magnet has two poles. A bar may be magnetized, however, so that it will have within itself a number of magnets, and hence a number of poles. The poles which occur in a magnet other than at the ends are called *consequent poles.* If the poles of a permanent magnet be drawn along a piece of steel (knitting needle) skipping at several places, Fig. 268, the needle will contain two or more consequent poles. It is possible to magnetize a steel bar in such a way that both ends may be either N-poles or S-poles. The bar will, of course, in this case contain consequent poles opposite in sign to those at the end.

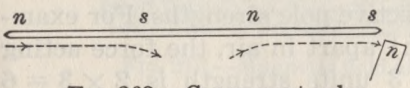


FIG. 268.—Consequent poles

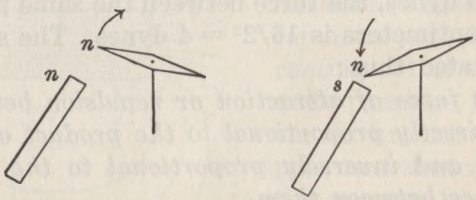


FIG. 269.—Like poles repel; unlike poles attract

304. The Laws of Magnetic Attraction and Repulsion. (a) Present to the north-seeking pole of the magnetic needle the N-pole of a magnet, Fig. 269, the like poles repel. Now present to the north-seeking pole of the needle the S-pole of the magnet; the unlike poles attract. These facts of attraction and repulsion are stated in the first law as follows:

I. Like poles repel; unlike poles attract.

(b) The pole strength of a magnet is measured by the force in dynes with which it attracts or repels a magnetic pole of unit strength. A *unit magnetic pole* is one which at a distance of one centimeter in air from an equal and similar pole repels it with a force of one dyne. The force of attraction or repulsion between two magnetic poles in a given medium depends upon the strength of the poles, and upon the distance which they are apart. First, the force of attraction or repulsion is *directly* proportional to the product of the respective pole strengths. For example at unit distance (1 cm) apart in air, the force acting between poles of 2 and 3 units strength is $2 \times 3 = 6$ dynes; and again, the force between poles of 4 and 5 units respectively is $4 \times 5 = 20$ dynes. Second, for given poles, the force of attraction or repulsion is *inversely* proportional to the square of the distance between them. That is, if the force acting between poles at unit distance apart is 16 dynes, the force between the same poles when placed 2 centimeters is $16/2^2 = 4$ dynes. The second law may be stated thus:

II. The force of attraction or repulsion between two poles is directly proportional to the product of the pole strengths, and inversely proportional to the square of the distance between them.

(c) The force of attraction or repulsion depends upon the nature of the medium in which the poles are placed. Thus, the magnetic attraction or repulsion of two given poles is slightly greater in a vacuum than in air.

III. The force of attraction or repulsion depends upon the nature of the medium in which the poles are immersed.

305. Magnetic Field and Lines of Force. The magnetic field is that portion of space surrounding a magnet which is affected by the magnet. The magnetic field may be mapped out by sprinkling fine iron filings upon a glass

plate and placing the plate over a magnet. Tap the plate gently. The filings arrange themselves in curved lines,

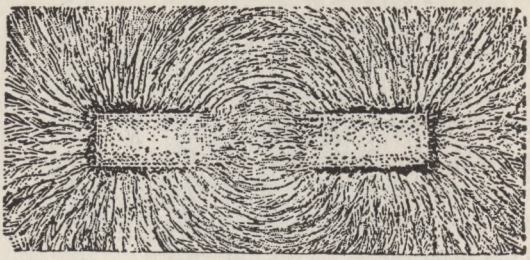


FIG. 270.—Magnetic field about a bar magnet

in response to the magnetic force acting upon them, Fig. 270. These lines were called by Faraday *lines of force;* they are sometimes called *lines of induction.* These lines of force are conceived of as coming out of the N-pole and passing in closed curves around to the S-pole, and thence through the magnet back to the N-pole, as shown in Fig. 271.

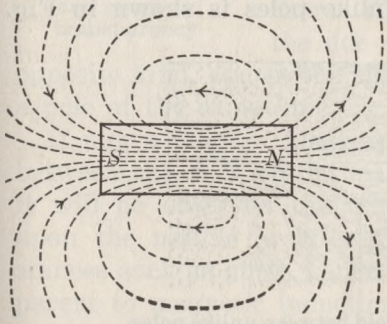


FIG. 271.—Lines of force come out the N-pole and enter the S-pole

The position which a magnetic needle assumes is determined by the direction of the lines of force of the field in which it is placed. When placed at a given point in a magnetic field, a magnetic needle always tends to take such a position that the lines of force of the field enter the S-pole of the needle and leave at the N-pole.

306. Magnetic Fields between Like and Unlike Poles. (a) In Fig. 272 there is shown the condition of the lines of force in a magnetic field between like poles. The

lines of force emanating from the respective N-poles are crowded together in the center of the field. This

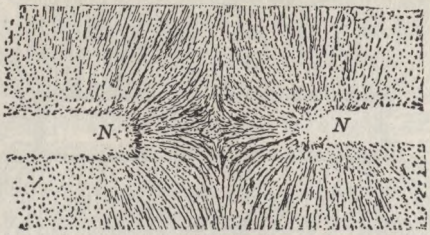


FIG. 272.—Magnetic field between like poles

crowding together of the lines suggests an explanation of the fact that like poles repel.

(b) The field between unlike poles is shown in Fig.

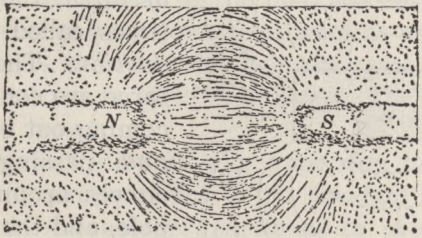


FIG. 273.—Magnetic field between unlike poles

273. Here the lines of force are represented as leaving the N-pole and entering the S-pole by more or less curved lines. These curved lines of force tend to contract, thus producing the phenomenon of attraction.

307. Magnetic Induction. We have seen that when a piece of iron is placed in contact with a magnet, it, too, becomes magnetized. The iron may be magnetized, however, without actually coming in actual contact with the magnet. Place one end of a soft iron nail in iron filings, and then bring near to the other end of the nail the pole

of a bar magnet. If now both nail and magnet be lifted, it will be observed that the tufts of iron cling to the lower end of the nail, Fig. 274, thus showing that it is magnetized even though not in actual contact with the permanent magnet. When the magnet is removed, the nail at once loses its magnetism, the iron filing falling away. While the nail is influenced by the presence of the magnet, it is said to be magnetized by *induction*. If the pole of a magnet be brought near a piece of iron or other magnetic substance, there will be induced in the iron on the side next the magnet a pole of the opposite kind, and in the side farthest from the magnet, a pole of the same kind.

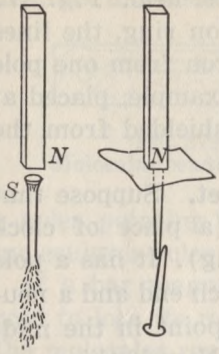


FIG. 274 FIG. 275
Magnetic induction and magnetic transparency

If now a piece of glass or a sheet of paper be slipped between the ends of the magnet and the nail, Fig. 275, it will be observed that the magnetic inductive effect upon the nail is practically undiminished. Many substances such as glass, paper, and wood seem to be transparent to magnetic induction. Such substances are said to be *magnetically transparent*.

308. Magnetic Shields. We have learned that glass is magnetically transparent. Iron, on the other hand, is not magnetically transparent in the sense in which glass is. This fact, together with the magnetic shielding effect of iron, is very well shown in Figs. 276 and 277. In Fig. 276 there is shown

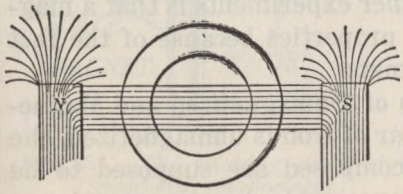


FIG. 276.—Magnetically transparent glass ring

the passage of the lines of force through the sides of a glass vessel placed in a strong magnetic field. Fig. 277 illustrates the shielding effect of an iron ring, the lines of force passing around through the iron from one pole to the other. An object, a watch for example, placed at *A* within the ring will be effectually shielded from the magnetizing force of the field.

309. The Effect of Breaking a Magnet. Suppose that we magnetize a thin strip of steel (a piece of clock spring). It has a pole at each end and a neutral point in the middle. Now if this magnet be broken at the middle, it will be found that each piece is a magnet, Fig. 278, having a pole at each end. If each piece be again broken, the smaller pieces will still be magnets. Thus we may conceive the breaking process to go on until the molecule is reached. The conclusion from this and other experiments is that a magnet possesses its magnetic properties because of the fact that its molecules are magnets.

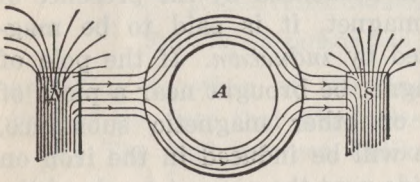


FIG. 277.—Magnetic shielding effect of a metal ring

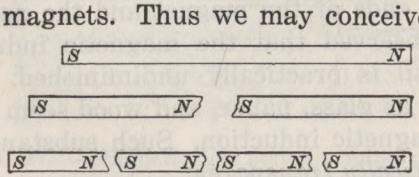


FIG. 278.—Effect of breaking a magnet

310. Molecular Condition of Unmagnetized and Magnetized Iron Bars. When a bar of iron is unmagnetized, the molecules of which it is composed are supposed to lie with their poles in every conceivable direction, as shown in Fig. 279. When the bar is magnetized, the molecules

are assumed to lie with their magnetic axes all in a definite direction, Fig. 280, that is, all the molecular

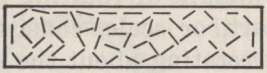


FIG. 279

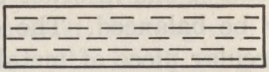


FIG. 280

Molecular condition of unmagnetized and magnetized bars

n-poles pointing toward the *N*-pole of the bar and the molecular *s*-poles toward the *S*-pole.

If a bar magnet be dropped or jarred violently, it is said to lose its magnetism. What really happens is that the molecules rearrange themselves again, taking up the positions as shown in Fig. 279. Strictly speaking, the bar has as much magnetism as before; what it has lost is its polarity.

311. The Magnetic Speedometer. Nearly all our common uses of magnetism are in connection with electric devices, and will be discussed when those are taken up.

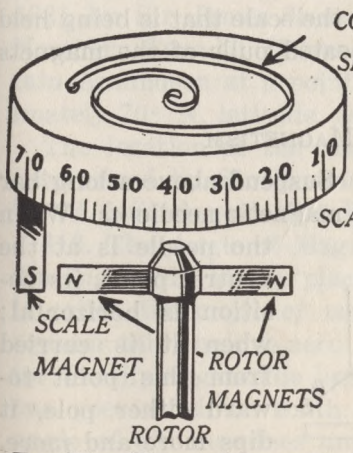


FIG. 281.—The working parts of a magnetic speedometer

The one common use of magnetism without electricity is the automobile speedometer. The mileage-indicating parts of this device are not magnetic in action and will not be discussed here. The speed-indicating portion, shown in diagram in Fig. 281, consists of a moving part, the

rotor, carrying two magnets with the same pole of each furthest away from the center. Two magnets are used instead of one in order that the weight of rotating parts may be balanced. The other part, or scale, consists of a circular piece of metal, freely rotating about the same center as the rotor, but held from continuous motion by a spring. On its outside it carries the scale of miles per hour, which passes beneath a pointer under a window in the casing of the machine. Beneath the scale is mounted a single magnet, with the opposite pole to those attached to the rotor.

As the rotor is driven by a flexible shaft connected to the drive-shaft of the automobile, the magnets on it exert a constant pull as they pass the scale magnet. This pull is repeated at more frequent intervals as the rotor is driven faster with the greater speed of the car. By adjusting the spring which holds the scale from revolving, the instrument can be set so that the exact speed of the car is indicated by the line on the scale that is being held under the pointer by the repeated pulls of the magnets on the rotor.

Terrestrial Magnetism

312. The Earth a Magnet. Suspend above a long bar magnet *NS*, Fig. 282, a short magnetic needle *ns*. When the needle is at the neutral point, its position is horizontal; when it is carried from this point toward either pole, it dips more and more, until the pole is reached, when its position is vertical. Now if a magnetic dipping needle, Fig. 283, be carried on the surface of the

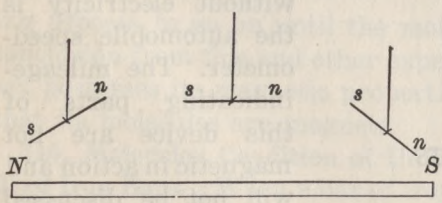


Fig. 282.—The earth a magnet

earth from the equator toward the geographic poles, it will be observed to dip in a similar manner. This means that the earth behaves exactly as if it were a magnet, having one magnetic pole near the north geographic pole and the other near the south geographic pole. The earth's magnetic pole at the north corresponds to the S-pole of a bar magnet; the magnetic pole at the south corresponds to the N-pole of a bar magnet. Since unlike poles attract, we thus understand why the N-pole of a needle points toward the north.

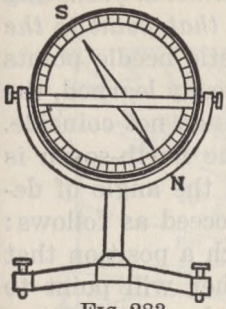


FIG. 283
Dipping needle

The magnetic poles of the earth do not coincide with the geographic poles. The magnetic pole of the north, for example, is more than 1000 miles from the geographic pole. The magnetic pole of the north was first located in 1831 by Sir James Ross at a point in Boothia Felix, Northern Canada. It was located again in 1905 by Captain Amundsen at a point a little farther west, approximately 70° N. latitude, and 97° W. longitude.

The location of the magnetic pole of the south was determined by Sir Ernest Shackelton at a point about 72° S. latitude and 155° E. longitude.

313. The Angle of Magnetic Dip. The angle which a dipping needle at any place makes with a horizontal line is called the *angle of magnetic dip.* At the magnetic equator this angle is zero; that is, at the magnetic equator the dipping needle lies in a horizontal direction. As one goes north or south of the magnetic equator, the angle of dip increases until at the magnetic poles it is 90°. At the north magnetic pole, for example, the dipping needle lies in a vertical direction. In the latitude of Ann Arbor, Michigan, the angle of dip is about 72°.

314. The Angle of Magnetic Declination. We sometimes say that the magnetic needle points north-south. This, however, is very rarely true. *True north from any point on the earth is the direction from that point to the north geographic pole.* Now the magnetic needle points to the magnetic pole and as we have already learned, the geographic poles and the magnetic poles do not coincide. The angle made by the needle and true north-south is called the *angle of declination.* To find the angle of declination for a given place one may proceed as follows: Drive two stakes into the ground in such a position that a string stretched from one to the other will point to the North star, which determines very closely the direction of the north geographic pole. The string therefore lies on a geograpic meridian; that is, it points true north-south. Now place a long magnetic needle just below the string. The angle made by the needle and the string is the angle of declination. Columbus used a compass on his first voyage to America. Up to this time it was supposed by navigators that while the suspended magnet in the compass did not always point north and south, it did always point in the same direction; that is, they supposed the angle between the geographical and magnetic meridians to be constant. Much to his own astonishment, and much to the dismay of his sailors, Columbus found that as he proceeded west, the needle pointed in a direction more and more to the east. He *thus discovered that the magnetic declination is not a constant but varies from one point to another on the earth's surface.*

315. Isogonic Lines. Lines drawn through places that have the same magnetic declination are called *isogonic lines.* In Fig. 284 there is a map showing the isogonic lines for the United States. The heavy black line is the line of *no declination.* Starting at the north magnetic pole, it runs through the western part of Hudson Bay,

Michigan, Ohio, eastern Kentucky, Tennessee, South Carolina, and thence through South America to the south magnetic pole. At all places located on this line the mag-

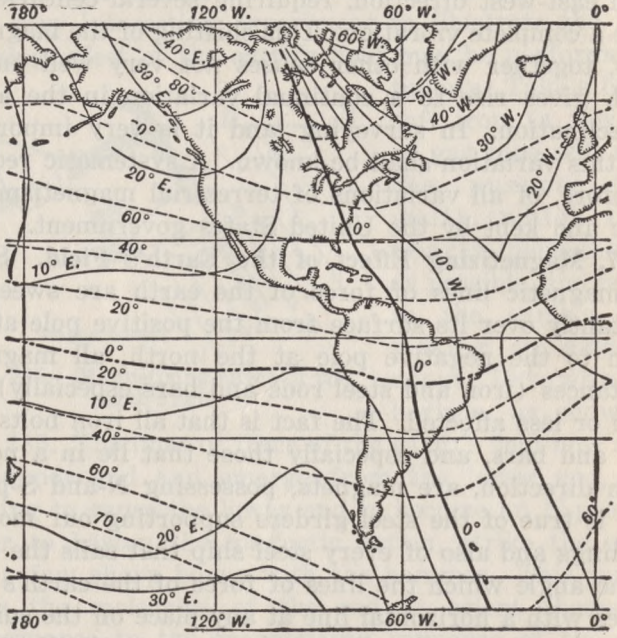


FIG. 284.—Line of no declination, Western Hemisphere

netic needle points true north; at all places east of this line the compass points west of north; at places west of this line it points east of north. The extreme variations in the United States are found in Maine and in Washington. In northeastern Main the compass points 20 degrees west of north; in the Puget Sound region it points nearly 25 degrees east of north.

The student should bear in mind that the direction of an isogonic line through any place is *not* the direction of the compass needle; it simply indicates the localities that have the same magnetic declination.

316. Variation in the Direction of the Earth's Field. The magnetic poles are not fixed in position. The north magnetic pole, for example, swings slowly back and forth in an east-west direction, requiring several centuries to make a complete vibration. This shifting of the magnetic poles, together with other causes not very well understood, gives rise to a continual variation in the angle of declination. In surveying land it is very important that this variation shall be known. A systematic record, therefore, of all variations of terrestrial magnetism are made and kept by the United States government.

317. Magnetizing Effect of the Earth's Field. Since the magnetic lines of force of the earth are sweeping constantly over its surface from the positive pole at the south to the negative pole at the north, all magnetic substances (iron and steel rods and bars especially) are more or less affected. The fact is that all iron bolts and rods and bars, and especially those that lie in a north-south direction, are magnets, possessing N-and S-poles. This is true of the steel girders supporting our modern buildings and also of every steel ship that sails the seas.

The angle which the lines of force of the earth's field makes with a horizontal line at any place on the surface of the earth is called the *angle of magnetic dip.* This angle, which may be determined by a dipping needle, is for places in the latitude of Michigan about 70°. This being the fact, it follows that iron bars, posts, telephone poles, flagstaffs, and other vertical objects which are made of iron or steel are magnetized in such a way as to cause the upper end (higher end) in each case to be an S-pole, and the lower end to be an N-pole; that is, the lines of force of the earth enter the upper end and come out the lower end. That this is true may easily be verified by testing the lower end of a steel post or rod with a small magnetic needle, such, for example, as that con-

tained in a small compass. It will be found that the lower end of the metal post or rod repels the N-pole of the magnetic needle.

A very interesting experiment to illustrate the magnetizing effect of the earth's field may be performed as follows: Take a bar of soft iron (a few feet of gas pipe will serve the purpose very well), which is as nearly demagnetized as possible, and hold it in the earth's field, making an angle of about 70° with the horizon, as shown in Fig. 285. The lines of force of the earth's field now enter the upper end and emerge from the lower end, thus tending to cause the lower end to become an N-pole. In order to hasten the magnetic action, strike the upper end a few sharp blows with the hammer. This jarring helps the molecules of the rod to orient themselves in response to the magnetizing influence of the field. Test the lower end of the rod. It will be found to be strongly magnetic, repelling the N-pole of the testing needle. Now reverse the rod and again strike the end with a hammer. The polarity of the rod is reversed; that is, the end which was an S-pole when placed downward in the field and subjected to the blows of the hammer soon becomes an N-pole.

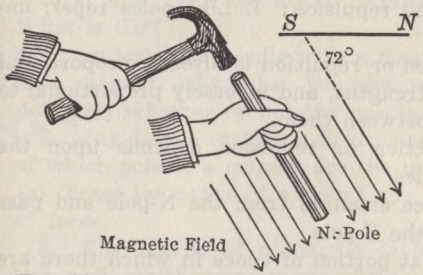


FIG. 285.—Magnetizing effect of the earth's field

In working in the laboratory with sensitive magnetic instruments great care must be exercised to guard against the magnetizing effect of adjacent magnetic substances, and account must also always be taken of the earth's field.

SUMMARY

1. Characteristic properties of the magnet: (a) magnetic attraction or repulsion; (b) polarity; (c) orientation.

2. A unit magnetic pole is one that at a distance of one centimeter in air from an equal and similar pole repels it with a force of one dyne.

3. Laws of attraction and repulsion: I. Like poles repel; unlike poles attract.

II. The force of attraction or repulsion is directly proportional to the product of the pole strengths, and inversely proportional to the square of the distance between them.

III. The force of attraction or repulsion depends upon the medium surrounding the poles.

4. Magnetic lines of force emanate from the N-pole and pass in closed curves around to the S-pole.

5. A magnetic field is that portion of space in which there are magnetic lines of force.

6. The angle of magnetic dip is the angle which a dipping needle at any place makes with the horizontal.

7. The angle of declination is the angle the compass needle at any place makes with a true north-south direction.

8. Isogonic lines are lines drawn through places having the same angles of declination.

REVIEW EXERCISES

A

1. (a) What is the distinction between a magnet and a magnetic substance? (b) Name several magnetic substances. (c) From the following list select the metals that are non-magnetic: steel, lead, silver, brass, nickel, gold.

2. A steel knitting needle is drawn several times from end to end across the S-pole of a bar magnet. (a) Which is the N-pole of the needle thus magnetized? (b) Why should the needle be drawn across the pole of the bar magnet always in the same direction?

3. Name the characteristic properties of a magnet, and give illustrations of each.

4. State the laws of magnetic attraction and repulsion.

5. Given two small cylindrical pieces of iron, *AB* and *CD*. One is soft iron and the other is a permanent magnet. By means of

a bar magnet you are required to determine which of the pieces (AB or CD) is a magnet? (a) When the N-pole of the bar magnet is presented to the end B of the first piece, attraction occurs; likewise when the S-pole is presented to B, attraction again occurs. What is the piece AB (magnetic substance or magnet)? (b) When the N-pole is presented to the end D of the second piece, attraction occurs; and when the S-pole is presented to D, repulsion occurs. What is CD?

6. In testing a needle with a magnet to determine if the needle is a magnet, why must you get repulsion before you can say definitely whether or not the needle is a magnet?

7. (a) Define magnetic field; magnetic line of force. (b) Out of which pole of a magnet are the lines of force assumed to come? (c) Trace by means of a sketch the complete path of a given line of force.

8. The N-pole of a magnet is brought near the head of an iron nail, which acquires magnetic poles by induction. Which end of the nail is an N-pole, the head or the point?

9. Make a sketch to illustrate the magnetic field between (a) like poles; (b) unlike poles.

10. (a) What is magnetic transparency? Name some substances that are magnetically transparent. (b) Explain how a ring of soft iron may be used as a magnetic shield.

11. (a) When a bar of iron is magnetized, how are its molecules supposed to be affected? (b) Why does heating or dropping a magnet tend to diminish its polarity?

12. (a) What evidence have we to prove that the earth is a magnet? (b) Where is its N-pole? its S-pole? (c) Are the magnetic poles stationary?

B

13. What is a unit magnetic pole?

14. Two magnetic poles of like kind, pole strength 2 and 20 units respectively, are placed 1 cm apart in air. (a) With what force in dynes do they repel each other? (b) What will be the force of repulsion when they are 4 cm apart?

15. Two like magnetic poles, m and m', 2 cm apart, repel each other with a force of 4 dynes. With what force will they repel each other when they are 4 cm apart?

16. Suppose that the pole strength of m (Ex. 15) be doubled

and the pole strength of m' be quadrupled. With what force will they now repel each other when they are 2 cm apart?

17. An iron bar is left lying for some time in a north-south direction. How will the north end of this bar affect the N-pole of a compass needle?

18. The S-pole of a magnetic needle when brought near the top of a steam radiator is repelled. Explain.

19. What is the effect of near-by deposits of iron ore, iron posts, wire fences, and the like upon the surveyor's compass?

20. What is magnetic dip? What is the value of the angle of dip (a) at the magnetic equator? (b) at the magnetic pole?

21. What is the distinction between *angle of dip* and *angle of declination*?

C

22. What are isogonic lines, and what do they tell us about angles of declination?

23. What is the line of no declination? If a person is east of this line, how will a compass needle point with reference to the north geographic pole? What is "true north" from a given place?

CHAPTER X

STATIC ELECTRICITY

ELECTRIFICATION

318. Introductory. Phenomena associated with electrical discharges, as seen in the flash of lightning, have always been a part of man's experience, and in modern times the use of electricity in the production of light, the ringing of door bells, and the running of trolley cars has become so familiar as to excite no more wonder than the phenomena associated with gravitation, heat, or sound. There is one very important difference, however, between electrical phenomena and those of heat or sound. Take heat for example. We know not only what it does, but we also know what it is. In the case of electricity, on the other hand, we know what it does but we do not know exactly what it is. We know that electricity may reside upon a body in the form of a "static charge," and that it will "flow" from one body to another producing a "current"; we know that the energy of a current of electricity may be transformed into heat, light, or mechanical motion, but just what electricity is no one at present is absolutely certain.

Electricity is apparently neither matter nor energy; it may be combined or associated with matter, and energy may be expended in moving it. Indeed its great importance to mankind lies in the fact that energy expended upon it in one part of a system, as in the electric power plant, may be made to reappear in another part of the system in the form of light, heat, work, or chemical change.

319. Electrification. The term electrification may be used in two entirely different senses, namely, (a) as the act of giving a body an electric charge, and (b) the condition of the body after being charged. Since the days of Thales, a Greek philosopher, who lived 640-546, B. C., it has been known that when certain bodies were rubbed together they possessed the power of attracting other bodies. For example, if a glass rod is rubbed with silk, it will attract a pith ball or bits of paper, Fig. 286. We say that the glass rod is *electrified.* In a like manner we may electrify a rod of sealing wax or vulcanite by rubbing it with flannel or cat's fur. Other familiar illustrations are seen in the electrification of a rubber comb when drawn through the hair, or in the electric discharge which occurs when the hand is drawn over a cat's back on a dry day. The substance which was earliest found to have this property of becoming electrified is amber, a kind of fossil resin, found on the shores of Greece. The Greeks called this amber *electron.* Later an English physicist, Gilbert (1540-1603), called the phenomenon *electric.* The word *electricity* came into use first in 1646.

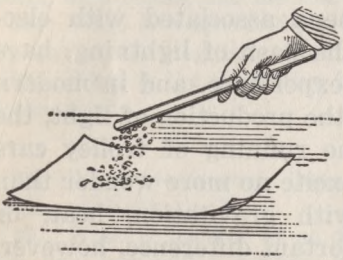


FIG. 286.—Bits of paper attracted by an electrified rod

Static electricity is electricity in a state of rest; *current electricity* is electricity in motion.

320. Two Kinds of Electrification. It was early discovered that there are two kinds of electricity, one called positive (+) and the other negative (—). These terms (+ and —) were suggested by Benjamin Franklin and have since been adopted the world over. *A glass rod rubbed with silk is said to be positively electrified; a rod*

of sealing wax or vulcanite rubbed with flannel is negatively electrified. There are many other substances which may be electrified, some positively and some negatively. We shall, however, for the sake of simplicity and clearness, think of glass rubbed with silk as a type of positive electrification, and sealing wax or vulcanite rubbed with flannel or cat's fur as a type of negative electrification.

There appears to be no very good reason for calling the electrical charge on glass positive, or that on sealing wax or vulcanite negative, other than the fact that these terms were adopted when the subject was first studied. Since sealing wax is very brittle and therefore is easily broken, vulcanite is usually used in the laboratory for producing negative charges.

321. The Electron Theory. The electron theory maintains that electricity has an atomic structure, and that each atom of which matter is composed is made up of two distinct parts, (a) a positively charged nucleus and (b) one or more electrons rotating around the nucleus somewhat as the planets of our solar system rotate around the sun. An *electron* is the elementary unit of negative electricity; it is the smallest particle known to science. All electrons carry negative charges of the same amount. In each atom there are under ordinary conditions just a sufficient number of electrons to neutralize the positive charge on the nucleus. This is why matter under ordinary conditions shows no sign of an electric charge.

The nucleus of the atom always carries a positive charge. The elementary unit of positive electricity on the nucleus is sometimes called the *proton.* The hydrogen atom carries one elementary positive charge in the nucleus, and around this nucleus there rotates one electron. The helium atom has two positive charges in the nucleus and two electrons. Electrons are free to move;

protons are firmly fixed in the nucleus and are not free to leave it.

322. Electrons and Electrification. The most modern theory of electrification is an application of the electron theory. We have just learned that under ordinary conditions there are in every atom just enough electrons to neutralize the positive charge on the nucleus. In this condition the atom, and consequently the matter of which it forms a part, is neutral; that is, it shows no sign of either positive or negative charge. But when, through friction or otherwise, some of the electrons are pulled out of the atom, there is left an excess of positive electricity and the atom is positively charged in virtue of its loss of one or more electrons. On the other hand, the body which received the electrons which were pulled away from the atom becomes negatively charged, in virtue of its excess of electrons. *All electrons bear negative charges.* They are quite easily dissociated from the influence of the positive nucleus of the atom. When they are thus freed, they may pass readily from one body to another. *An excess of electrons on a body means a negative charge; a deficiency of electrons, a positive charge.*

Now when a glass rod is rubbed with silk, the friction of the silk against the rod pulls electrons from the glass. The silk has gained electrons, therefore it is negatively electrified; the glass rod, having lost electrons, is positively electrified. The charge gained by the silk is just equal to the charge lost by the glass. Also, when a vulcanite rod is rubbed with flannel, electrons pass from the flannel to the vulcanite; the vulcanite therefore becomes negatively charged. The negative charge acquired by the vulcanite is exactly equal to the positive charge on the flannel. Whenever a charge of one sign appears on one body, there is always a corresponding charge of the opposite sign on some other body.

It thus appears that electrification is a process which is associated with the transfer of electrons from one body to another. When electrons are given to a neutral body, it becomes negatively charged; when electrons are taken away, it becomes positively charged.

Attraction and Repulsion

323. The Laws of Electrostatic Attraction and Repulsion. Among the most familiar phenomena of electrification are those of attraction and repulsion. If a rod be electrified and brought near a light body, such as a pith ball suspended by means of a fine silk thread, two things may be observed: (a) The pith ball is at first attracted to the electrified rod, and (b) after being in contact for a moment it is repelled. It is important to note that when an electrified body is brought near another body, attraction or repulsion always results.

The first law of attraction and repulsion may be demonstrated experimentally as follows: Charge a glass rod

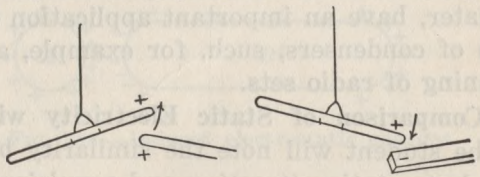


Fig. 287.—Charges of like signs repel; charges of unlike signs attract

positively and suspend it in a stirrup, as shown in Fig. 287. (a) Bring near to the suspended positively charged rod another glass rod also positively charged; the two repel each other. (b) Now present to the suspended positively charged glass rod a vulcanite rod which is negatively charged; the two attract each other. The laws of attraction and repulsion may be stated as follows:

1. *Charges of like signs repel; charges of unlike signs attract.*

II. *The force of attraction or repulsion between two charges varies directly as the product of the charges,* and *inversely as the square of the distance between them.* A *unit charge* of static electricity is that quantity which at a distance of one centimeter in air from an equal charge of the same sign repels it with a force of one dyne.

III. *The force of attraction or repulsion between two charges depends upon the nature of the medium in which the charged bodies are placed.* For example, the force of attraction between charges of unlike signs at a given distance apart in air is greater than when the medium between the charged bodies is paraffin or glass. The medium separating the charges is called the *dielectric.* For example, suppose two unlike charges are at a distance of one centimeter apart. When the medium separating the charges is air, the force of attraction will be 2 times as great as when the dielectric is paraffin, and 8 times as great as when glass is used.

The laws of attraction and repulsion, as will be explained later, have an important application in the construction of condensers, such, for example, as are used in the tuning of radio sets.

324. Comparison of Static Electricity with Magnetism. The student will note the similarity between the laws of electrostatic attraction and repulsion and those of magnetism, as explained in Art. 304. Both in electrostatics and in magnetism, like signs repel and unlike signs attract. Also, the forces of attraction and repulsion vary inversely as the square of the distance between the magnetic poles or between the charged bodies, and these forces also are affected by the nature of the medium between the poles or the charged bodies.

Because the laws of attraction and repulsion are the same in both cases, it must not be assumed that the phenomena of electrostatics are similar in all respects

to those of magnetism; indeed, in certain particulars quite the contrary is true. (a) In electricity we have two charged particles, the negatively charged electron which is free to move, and the positively charged proton which is stationary in the nucleus of the atom. In magnetism there are no such negative and positive magnetic particles which are capable of being separated from each other. (b) In electrostatics, we have no such thing as permanent positive and negative poles as is the case in magnetism. It is true that if a charged body

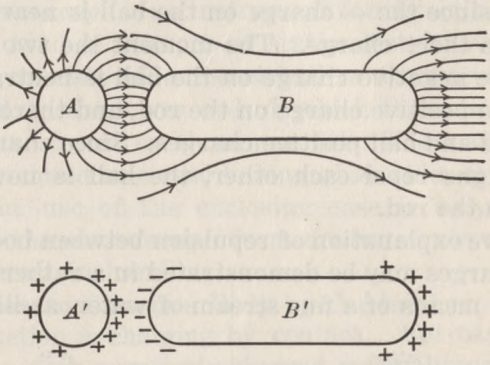


FIG. 288.—Induced electrostatic charges

A' is brought near the uncharged rod B', Fig. 288, the rod will be charged by induction, one end being negatively charged and the other positively charged; but the moment A' is removed, the negative and positive charges on B' will neutralize each other and as a consequence B' will be neutral.

325. Discussion of Attraction and Repulsion. We are now prepared to explain, in a measure, the phenomena of attraction and repulsion. If a glass rod positively

charged be brought near a pith ball, Fig. 289, there will be induced on the side of the pith ball nearest the rod a — charge, and on the side farthest from the rod a + charge. The induced positive charge on the ball is exactly equal in quantity to the induced negative charge. Now since unlike signs attract and like signs repel, it follows that the — charge will tend to move toward the glass rod and the + charge away. The force of attraction, however, is greater than the force of repulsion, since the — charge on the ball is nearer to the rod than is the + charge. The moment the two come in contact, the negative charge on the ball is neutralized by some of the positive charge on the rod, and there remains on both rod and ball positive charges. Since charges having like signs repel each other, the ball is now driven away from the rod.

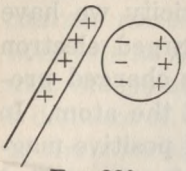


FIG. 289
Attraction and repulsion

The above explanation of repulsion between bodies having like charges may be demonstrated in a rather striking manner by means of a fine stream of water, as illustrated

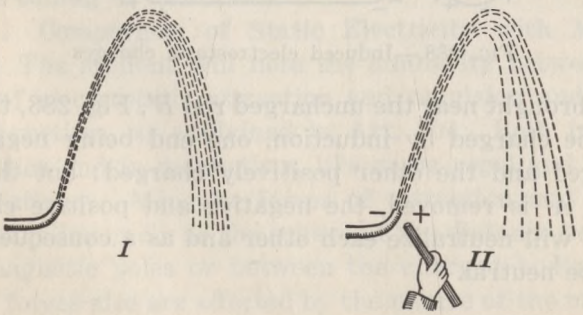


FIG. 290.—Electrical repulsion as illustrated by a spray of water

in Fig. 290. A fine stream which breaks up into a spray of droplets is made to issue from a nozzle. If a rod

charged positively (say) is brought near the stream, the spray widens perceptibly. This widening of the spray is due to the fact that the droplets of water have induced upon them negative charges, which repel each other.

Charging by Conduction and Induction

326. The Electroscope. A body may be charged electrically in two ways, (a) by *conduction* and (b) by *induction*. In order to demonstrate these two methods of charging it is necessary to make use of an electroscope.

An *electroscope* is an instrument for detecting the presence and kind of charges of static electricity. The electroscope consists of a metal ball or disc attached to a metal rod, to the lower end of which there is suspended two leaves of gold or aluminum foil. The leaves of gold foil are enclosed in a case having glass sides or in a glass flask. The use of the enclosing case is to protect the leaves of the electroscope from drafts of air or injuries from contact in handling.

327. How to Charge a Body by Conduction. Charging by conduction is charging by contact. For example, if a negatively charged rod be brought into actual contact with the knob of an electroscope, Fig. 291, some of the negative charge on the rod will flow off onto the electroscope, distributing itself over the knob and leaves.

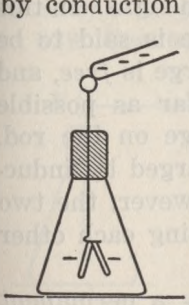


Fig. 291
Charging an electroscope by conduction

When the electroscope is charged by conduction, as shown in Fig. 291, the leaves being charged with electricity of like signs repel each other. When the charging body is removed, it will have lost a quantity of electricity equal to that gained by the electroscope. *When an electroscope is*

charged by conduction, the charge upon it is similar in sign to that on the charging body.

Sometimes it is desired to convey a definite portion of a charge to the electroscope. This may be done by means of a *proof plane,* which consists of a small metal disc attached to a nonconducting handle, Fig. 292, and which is used to transfer small charges from one body to another. For example, if a proof plane be touched to a charged body, such as the pole of an electric machine, and then conveyed to the knob of the electroscope, a slight divergence of the leaves will result. Upon conveying a second charge, a further divergence will occur. In this way it is possible to convey as much electricity to the electroscope as is desired.



FIG. 292
Proof plane

328. **How to Charge a Body by Induction.** Charging by induction is charging without bringing the bodies into actual contact. If a charged body be brought near another body, Fig. 293, there will appear upon the second body two charges, equal in quantity and opposite in signs. In this case the + charge is said to be *bound;* the — charge is *free,* and tends to get as far as possible from the — charge on the rod. The body *B* is charged by induction. If the charging rod be removed, however, the two charges on *B* will flow together, neutralizing each other and thus leaving *B* uncharged.

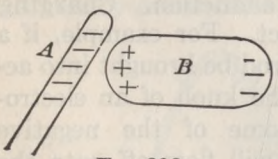


FIG. 293
Charging by induction

It is frequently desired to give a body a permanent charge by induction. Suppose, for example, we wish to charge an electroscope by induction. Proceed as shown in Fig. 294. Bring a charged body (negatively charged, say) near the knob of the electroscope. There will be

induced upon the knob a positive charge and upon the leaves an equal negative charge, *I*. The positive charge is bound, and the negative charge free. Now if a finger be touched to the electroscope, *II*, the free negative charge,

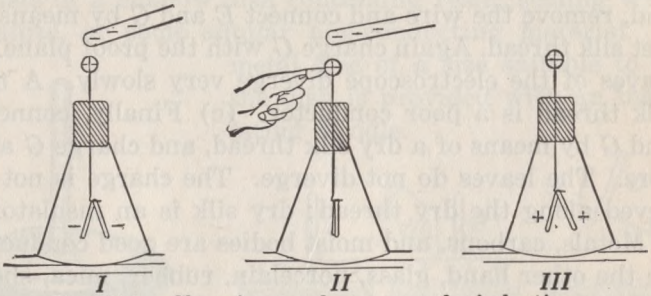


FIG. 294.—Charging an electroscope by induction

being repelled by the charge on the rod, will flow off to the earth. Remove the finger, and afterward remove the charging body. There will then remain upon the electroscope a positive charge which will cause the leaves of the electroscope to diverge, as in *III*. The electroscope is now charged by induction. *When a body is charged by induction, the charge which is imparted to it is of the opposite sign to that on the charging body.*

329. Conductors and Insulators. Those substances which conduct electricity readily are classified as *conductors;* those which do not conduct electricity readily are known as *nonconductors* or *insulators*. An experimental classification of substances as conductors or insulators may be made as follows: (a) Connect the knob of an electroscope *E* to an insulated metal globe *G*, Fig. 295, by means of a fine copper wire. Then by means of a proof plane convey a charge from the pole of an electric machine or

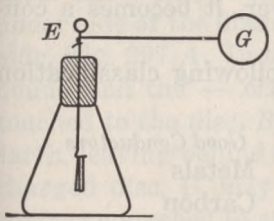


FIG. 295
Electrical conduction

other charged body to *G*. Instantly the leaves of the electroscope diverge, thus showing that a part of the charge imparted to *G* was conducted along the line to *E*. The copper wire is a good conductor of electricity. (b) Second, remove the wire and connect *E* and *G* by means of a wet silk thread. Again charge *G* with the proof plane. The leaves of the electroscope diverge very slowly. A moist silk thread is a poor conductor. (c) Finally, connect *E* and *G* by means of a dry silk thread, and charge *G* as before. The leaves do not diverge. The charge is not conveyed along the dry thread; dry silk is an insulator.

Metals, carbons, and moist bodies are good conductors; on the other hand, glass, porcelain, rubber, mica, shellac, dry silk, vaseline, and oils are insulators. It should be noted, however, that no hard and fast distinction can be drawn between conductors and non-conductors. There are no substances which are perfect conductors and likewise there are no substances which are perfect insulators. The terms *conductors* and *non-conductors* are purely relative. The conductivity of all substances depends upon their physical condition, especially with reference to temperature and moisture. Glass, for example, at ordinary room temperature is a non-conductor or insulator; when heated to a temperature of 200° C, however, it becomes a conductor.

Under ordinary conditions the following classification holds:

Insulators	*Poor Conductors*	*Good Conductors*
Dry air	Wood	Metals
Glass	Paper	Carbon
Porcelain	Alcohol	Water solutions of salts and acids
Rubber	Kerosene	
Paraffin	Pure Water	
Sulphur		
Resin		

330. The Electrophorus. It is sometimes desirable to charge a body with a larger quantity of electricity than can be obtained upon a glass rod or a vulcanite rod. To do this an *electrophorus* may be employed, Fig. 296. This consists of a shallow dish containing resin, sealing wax, vulcanite, or some similar non-conducting material. A metal disc of a size suitable to fit the dish is provided with an insulating handle.

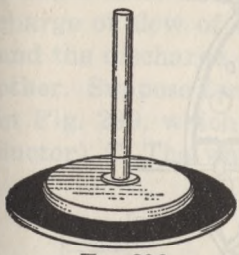


FIG. 296

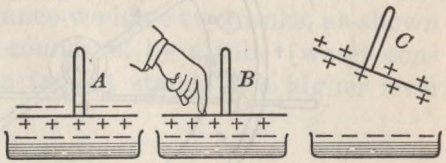


FIG. 297

Electrophorus, and diagram showing method of charging

To charge the electrophorus apparatus, we rub the non-conducting material with flannel or cat's fur, thus giving it a — charge. If now we place the metal disc upon the surface of the apparatus, Fig. 297, only a very small quantity of electricity will flow to the metal, because of the non-conducting property of the resin or other substance used. The — charge on the resin induces a + charge on the under side of the disc and an equal — charge on the upper side, Fig. 297 *A*. In this case the + charge on the disc is bound and the — charge is free. Now if the finger be touched to the disc, *B*, the free — charge will flow to the earth, leaving behind the bound + charge. The positively charged disc, *C*, may be removed and the charge on its surface conveyed to any other body.

331. The Electric Machine. An electric machine, shown in outline in Fig. 298, is nothing more than a device involving the principle of the electrophorus. It consists usually of one or more glass plates, which on being ro-

tated become charged, mainly by induction. A cross metal bar serves to draw off the free electricity from the rotating disc, and the bound charge is then transferred to the discharging points.

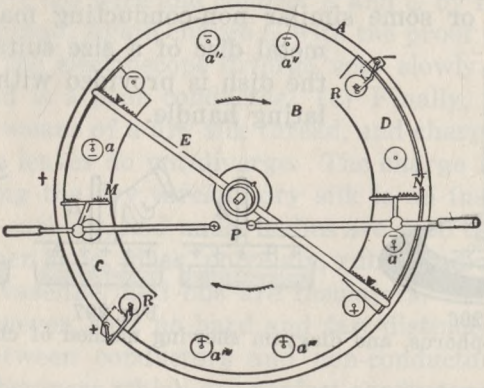


FIG. 298.—Essential parts of the Toepler-Holtz electric machine

When the electric machine is in operation, that is, when the movable glass disc is caused to rotate about its axis, the metal parts of the apparatus become highly charged; if these discharging knobs be brought close together, a bright spark occurs. Should contact between the charged Leyden jars of the machine be accidentally made by the hands of the operator, a severe shock will be received. To discharge the machine, bring the movable metal knobs into contact by means of the insulating handles.

Many interesting and striking experiments may be carried out by means of the electric machine.

POTENTIAL

332. Electric Pressure or Potential. We have learned that when a charged body A (a charged proof plane, for example) is brought in contact with an uncharged body

B, some of the charge on A is transferred to B. We have also learned in our experiments with the electric machine and the Leyden jar that when a conductor carrying a high positive charge is brought near a negatively charged conductor, the transfer of electricity from one to the other takes place violently in the form of a spark discharge. Now there is a close analogy between the discharge or flow of electricity from one body to the other and the discharge or flow of water from one tank to another. Suppose for instance we have two tanks, as shown in Fig. 299, which are connected by a tube (water conductor) T. The water in tank A stands at a higher level

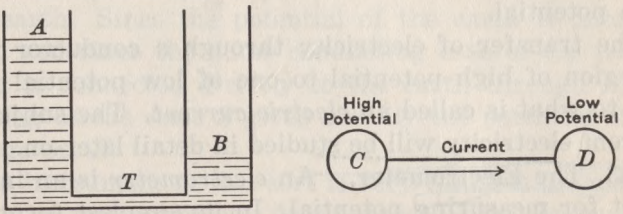


FIG. 299.—Analogy between hydrostatic pressure and electric potential

than that in tank B. The water pressure (hydrostatic pressure) in A is greater than that in B, consequently the water flows through the tube T from A to B; that is to say, the water flows from a region of high pressure to one of low pressure. Or, to look at it another way, if we find that the water flows from tank A to tank B, we are justified in saying that the hydrostatic pressure in A is higher than that in B. In a like manner if by means of a conductor (wire) we connect an electrically charged body C to a second body D, and a transfer (flow) of electricity takes place from C to D we say that the electrical pressure (potential) of C is higher than the electrical pressure or potential of D.

In this connection two things should be borne in mind by the student: First, electrical pressure (so called) is only *analogous* to hydrostatic pressure; in certain respects electrical pressure is quite different from water pressure. For example, pressure is measured in terms of force; potential is measured in terms of work. The difference of potential between two points is measured by the work required to carry a unit charge from one point to the other. In the second place, while the term "electric pressure" is quite commonly employed, the more precise scientific term is *potential. Electricity always tends to flow from a point of high potential to one of low potential.* We sometimes speak of the earth as having zero potential.

The transfer of electricity through a conductor from a region of high potential to one of low potential gives rise to what is called an *electric current.* The subject of current electricity will be studied in detail later on.

333. The Electrometer. An *electrometer* is an instrument for measuring potential. In its simplest form it is an electroscope calibrated to give readings in terms of units of potential difference. A very effective form of electrometer, or electrostatic voltmeter as it is sometimes called, is shown in Fig. 300. The binding post *A*, which is carefully insulated from the circular metal case, is connected with the metal rod *R*, which in turn supports at its lower end a light aluminum vane *L*, mounted so as to swing freely in a vertical plane about its middle point. To the lower end of *L* there is attached a fine pointer which moves over the scale *S*.

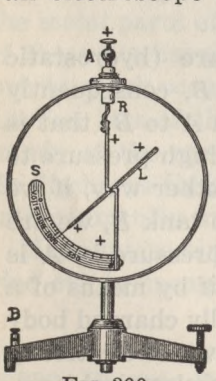


FIG. 300
Electrometer

Suppose that we wish to measure the potential of a

body X having a positive charge of electricity. We proceed thus: First, by means of a fine wire connect X with the binding post A. Then by means of a second wire connect the outer circular case of the instrument at B to the earth through the gas or water pipe system. Some of the charge on X flows through the conducting wire to A, and thence to the rod R and the vane L. The like charges on the vane and its support repel each other, thus causing L to swing to one side into a given position, depending on the potential of the charge on X. The greater the potential of X the greater will be the deflection of L. The deflection of the pointer over the scale S indicates directly the potential of X with reference to the earth. Since the potential of the earth is taken as zero, and since the outer conducting case of the instrument is connected directly to the earth through B, the reading of the scale S gives directly the potential of the charge on X.

An instrument of this sort is very valuable for making potential measurements, first because of its comparative simplicity, and second because of the range over which readings may be taken. For example, by its use potential readings may be made from as low as 0.001 volt to over 100,000 volts.

Other methods of measuring potential difference (voltage) will be discussed in a later chapter.

Electric Condensers

334. The Condenser. Under ordinary conditions when the quantity of electricity (charge) upon a body is increased, the potential of the charged body is correspondingly increased. It is often desirable, however, to increase the charge without increasing the potential. To do this an electric condenser is used. A *condenser* is a device for

increasing the charge without increasing the potential.

One of the earliest types of condensers used is the Leyden jar, which consists of a glass vessel coated about two-thirds of the way up on both the inside and the outside with tin foil, Fig. 301. A metal conductor communi-

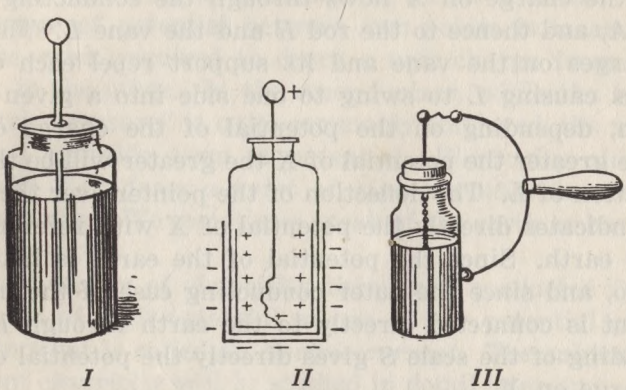


FIG. 301.—The Leyden jar

cates with the inner coating of the vessel. Experiments with this apparatus were first made at Leyden, Holland, in 1745, hence the name "Leyden" jar. To charge the jar we grasp the outer coating in the hand and place the knob in contact with one of the poles of an electric machine, thus charging the inner coating by conduction.

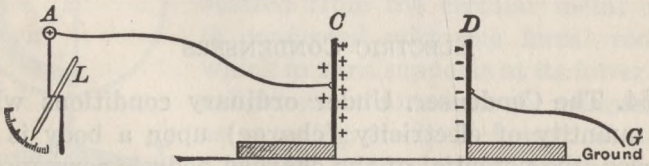


FIG. 302.—Testing the capacity of a condenser

The outer coating becomes charged through the glass by induction. The principle upon which a condenser works

may be demonstrated by a very simple apparatus, used in connection with an electrometer, Fig. 302. Fasten two metal conductors (sheets of tin, zinc, or copper about a foot square) to wooden supports. Place the conductors *C* and *D* face to face about a foot apart. By means of a fine wire connect the conductor *C* to the electrometer *A*, and the conductor *D* to the ground at G. (a) With a proof plane, charge *C* from an electric machine. The charge on *C* will induce a bound charge of the opposite sign on *D*, the free charge flowing through *G* to the earth. The vane *L* of the electrometer is deflected through a given angle, thus indicating the potential of the system. (b) Push *D* nearer to *C*, and it will be observed that the deflection of *L* diminishes, thus showing that the potential of the system has been lowered. (c) Now add small charges to *C* until the deflection of *L* is the same as first observed. This means that by diminishing the distance between *D* and *C* we are enabled to increase the charge on *C* without increasing the potential above its original value. (d) Repeat the operation, adding a sufficient charge to *C* each time to bring the potential reading of the electrometer up to its initial value.

It thus appears that by means of two conductors *C* and *D*, which constitute a condenser having air as the dielectric, it is possible by changing the distance between them, to vary the charge on the system without making a corresponding change in the potential.

The *electrical capacity* of a body is measured by the charge required to give it unit potential. The capacity of a condenser is increased (a) by increasing the number and size of the plates, (b) by bringing the plates nearer together, and (c) by using a substance having a greater dielectric constant.

335. Commercial Condensers. The modern commercial condenser consists of two or more sheets of tin foil sep-

arated by glass, mica, or paraffined paper, or other nonconducting substance, enclosed in a suitable case. The conventional method of representing a condenser in diagrammatic outline is illustrated in Fig. 303. Condensers are very largely used in the manufacture of induction coils, and in the operation of many kinds of modern electrical apparatus, such as the telephone and radio.

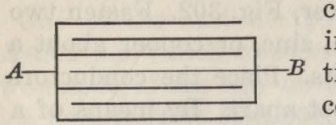


FIG. 303.—Diagram of a condenser

To meet the requirements of modern radio service a special type of variable condenser, such as shown in Fig. 304, is now manufactured in large quantities. The plates used in this condenser consist of sheets of aluminum. There are two sets, one fixed and the other movable about an axis *S*. The dielectric is air. We learned in Art. 334 that the capacity of a condenser may be varied by moving the plates nearer together or farther apart. In the variable air condenser change of capacity is secured by rotating the set of plates which are attached to the shaft *S*.

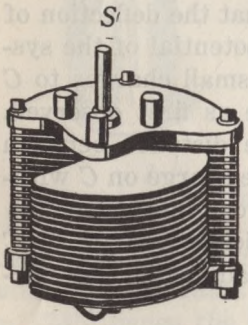


FIG. 304
Variable condenser

DISTRIBUTION OF AN ELECTRIC CHARGE

336. Static Charges Reside Upon the Surface. Place a tin cup or other hollow metallic vessel upon a nonconducting base of glass or paraffin and charge the cup from an electric machine. Touch a proof plane to the outside of the vessel and then to the knob of an electroscope; the leaves are affected, showing that there is a charge on the outside of the vessel. Now, touch the proof plane to the inside of the vessel and again to the knob of the electro-

scope; the leaves are not affected, showing that there is no charge on the inside. To prove that even for very high potentials the charge remains upon the surface, Faraday used a cage made of metal rods, inside of which he placed a sensitive electroscope. He then charged the cage so heavily that a spark discharge occurred from different points on the surface; the electroscope, however, was not affected in the slightest degree.

It has been found by many such experiments that a charge of electricity at rest (static electricity) always resides upon the surface of the conductor.

337. **Distribution of the Charge.** The distribution of the charge of static electricity upon the surface of a conductor depends upon two factors: first, upon the shape of the conductor, and second, upon the presence or absence of other charged bodies.

In the case of a spherical conductor free from the in-



FIG. 305

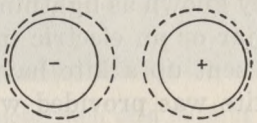


FIG. 306

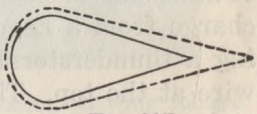


FIG. 307

Distribution of electrical charges

fluence of other bodies, the charge is distributed uniformly over the surface, Fig. 305. When there are present other charged bodies, the charge is distributed somewhat as shown in Fig. 306. In the case of an irregular shaped body, the charge has its greatest density at the pointed portion of the conductor, Fig. 307.

338. **Influence of Sharp Points.** Knowledge of the effect of sharp points in connection with conductors is of great importance in the operation of certain types of electrical apparatus. A sharp point upon a body serves to discharge the electricity, as may be shown by the fol-

lowing experiment. If an electroscope be charged, the leaves will remain apart for some considerable time, depending upon the condition of the atmosphere and the nature of the instrument. If now the head of a pin or other sharp pointed body fastened to the end of a glass rod be brought in contact with the knob of the electroscope, the leaves will collapse very quickly, thus showing the effect of the discharging point.

Small drops of moisture or particles of dust on electrical apparatus designed for use in electrostatics serve as discharging points. In order, therefore, to perform experiments with the electroscope, electric machine, or similar instruments, it is necessary that the apparatus be free from dust and that the experiments be performed when the air is dry.

339. Lightning. Benjamin Franklin was one of the first to demonstrate the similarity between the discharge of atmospheric electricity known as lightning and the discharge from a Leyden jar or an electric machine. During a thunderstorm he sent up a kite having a pointed wire at the top. The kite was provided with a hempen string, to the lower end of which was attached a key. In order to control the kite and at the same time to prevent the charge from passing through his body, Franklin fastened to the kite string a piece of nonconducting silk ribbon, the end of which he held in his hand. When the hempen string became wet he succeeded in drawing from the key electric sparks which he concluded were similar in every respect to those that came from the Leyden jar. This experiment is a dangerous one, and cases are on record where experimenters have lost their lives in trying to repeat it.

When two clouds oppositely charged come near to each other, the difference of potential (electric pressure) may be so great that discharge takes place, resulting in a flash

of lightning. The discharge from one cloud to another or from a cloud to the earth is similar in every respect to the discharge between the knobs of an electric machine, except that in the case of the atmospheric discharge the phenomenon occurs on a very much grander scale.

Thunder is the sound which accompanies a flash of lightning; it is caused by the sudden expansion of the air

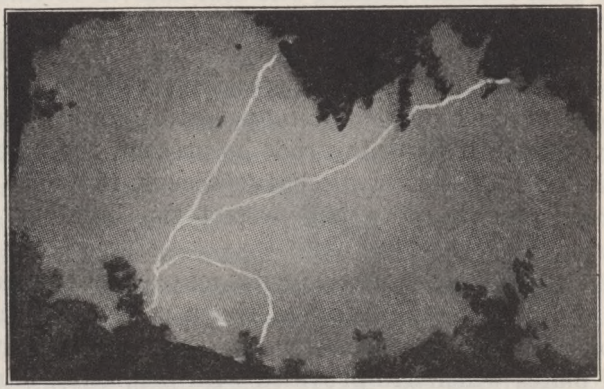


FIG. 308.—Photograph of a flash of lightning, showing foliage in the foreground and the illuminated sky in the background

due to the flash. When a flash of lightning occurs, the air along the line of discharge is heated, causing an increase in volume. This sudden expansion of the air gives rise to violent vibrations of the atmosphere, causing the crashing or rumbling noise called thunder. Thunder, therefore, always follows the flash of lightning and never precedes it.

In Fig. 308 there is given a reproduction of a photograph of a flash of lightning, showing foliage in the foreground and the illuminated heavens in the background.

340. The Lightning Rod. The lightning rod consists of a metal conductor provided with one or more sharp

points, and having the lower end connected with the earth. It is believed that the principal function of the rod is to furnish a means of silent discharge between the earth and the electrified atmosphere above. An electrified cloud induces a charge of opposite sign on the earth and objects below it, thus forming a condenser on an immense scale, in which the cloud corresponds to one plate and the earth to the other. The dielectric is the air between the cloud and the earth. If the building be provided with lightning rods, the sharp points furnish a means of silent discharge, thus tending to keep the electric pressure below the striking potential. In some cases, however, even if lightning rods be present, the potential may be so high that a discharge occurs, in which case the rods may or may not protect the building from damage.

A very instructive experiment may be performed to show the discharging effect of sharp points, which correspond to lightning rod points. Set up two circular metal plates which are provided with insulating handles, as shown in Fig. 309. The upper plate is held in place by means of a clamp which is fastened to the insulating handle. By means of wires, connect the plates to the knobs of an electric machine. Provide two metal standards with supporting metal bases, *P* and *B*, one having a knob at the top, the other having a sharp point. The upper plate corresponds to the cloud, the lower plate to the earth. The knob corresponds to a house or a tree, and the pointed rod to a lightning rod. First,

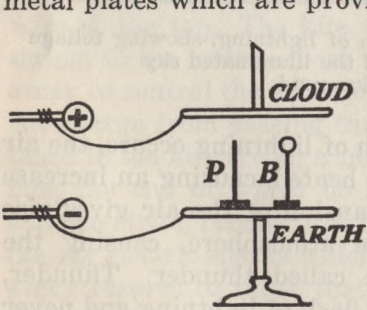


FIG. 309.—Experiment to illustrate lightning rod effects

place *B* alone between the plates. Start the machine and then lower the upper plate until the discharge distance is reached. Sparks pass freely from the upper plate to the knob. Now insert *P* and again operate the machine. No discharge occurs. If the point is removed, discharge between the plate and the knob again takes place.

341. Damage by Lightning. Damage done by lightning may be divided into two general classes: first, heat-effect damage, as, for example, where telephone or power conductors or other metal members lying in the path of the stroke are fused; second, mechanical disruptive disintegration, as, for example, the splitting of a tree, the breaking of a rock, or the overturning of a stack. Lightning seldom seriously damages such objects as railroad trains or locomotives, buildings having frameworks wholly of metal, grounded steel windmill towers, steel, battleships, and business blocks in cities. Apparently all of these objects conduct electricity sufficiently well for the electricity induced on them by the cloud to be dispersed by the point action before the difference of potential between the cloud and the object becomes great enough to produce a flash. If a lightning-stroke to such and object does occur, the mass of metal in it is usually sufficient to conduct the current safely to the ground. Conversely, lightning often damages nonconducting objects such as country homes and barns, school houses and churches, stacks, trees, cattle and horses, particularly where the stock is near wire fences. In general, it is very dangerous for a person to seek refuge under a tree during a thunder storm.

It is now known that a lightning rod need not be made of any particular sort of metal or that it be of any special shape. The important facts to be kept in mind in installing lightning rods are (a) that the wires are to be so put up as to furnish a number of discharging points at the

top of the building, and (b) that the lower ends of the wires be buried deep enough so as to be in contact with moist earth.

342. Electric Screens. It is believed that the most effective protection from lightning is furnished by metal screens. This may be demonstrated by a simple device involving the principle of Faraday's celebrated screen experiment. Within a wire screen place a small electroscope and then put the apparatus thus formed between the knobs of an electric machine, Fig. 310. When the machine is operated a discharge takes place across the screen, during which the electroscope remains undisturbed.

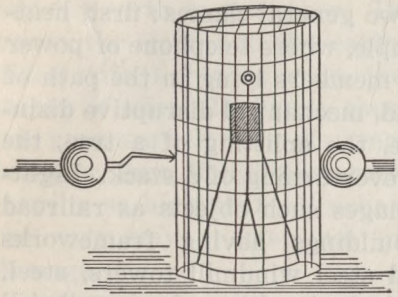


FIG. 310.—Electric screen

Warehouses and other buildings containing highly inflammable material are sometimes protected in this way by having stretched over them a network of metal conductors which are thoroughly grounded at a number of points. The screen effects of the metal roofs, girders, gas and water pipes which cover the modern structures of our cities probably explain why such buildings are so rarely injured by lightning.

SUMMARY

1. Positive electricity is the kind that appears on glass when rubbed with silk; negative electricity, the kind that appears on sealing wax or vulcanite when rubbed with flannel.

2. Every atom contains both kinds of electricity, a positive nucleus and negative electrons.

3. Electrification is a process of adding electrons to or taking

them from a body. An excess of electrons means a negative charge; a deficiency of electrons, a positive charge.

4. Laws of attraction and repulsion: I. Charges of like signs repel; charges of unlike signs attract.

II. The force of attraction and repulsion between two charges is directly proportional to the product of the charges and inversely proportional to the square of the distance between them.

III. The force of attraction or repulsion depends upon the nature of the dielectric. The dielectric constant of dry air is taken as 1.

5. Unit charge is that quantity of static electricity which at a distance of one centimeter in air from an equal charge of like sign repels it with a force of one dyne.

6. A body charged by conduction has a charge of the same sign as that of the charging body.

7. When a body is charged by induction, the charge remaining upon it is of the opposite sign to that of the charging body.

8. A conductor is a substance through which the electrons pass readily; a non-conductor or insulator is a substance through which the electrons cannot pass readily.

9. The difference of potential between two points is measured by the work required to carry a unit charge from one point to tne other. Electricity flows from points of high potential to those of low potential.

10. An electric current is a stream of electrons flowing from one point to another.

11. The electrical capacity of a body is measured by the quantity of electricity required to charge it to unit potential.

12. A condenser is a device for increasing the charge without increasing the potential.

13. The capacity of a condenser depends upon the size and number of the plates, their nearness together, and the dielectric.

14. Static charges of electricity are always upon the surface of the charged body. Sharp points facilitate the quiet discharge from the body to the air.

REVIEW EXERCISES

A

1. What is the distinction between static electricity and current electricity?

2. (a) What electric charge is developed on glass when it is rubbed with silk? (b) What charge is developed on vulcanite

when it is rubbed with flannel? (c) What is positive electricity? negative?

3. Discuss briefly the electrical constitution of the atom. In the neutral atom how do the negative charges of the electrons compare in number with the positive charges of the nucleus?

4. According to the electron theory, when is a body (a) negatively charged? (b) positively charged? (c) neutral?

5. State the laws of electrostatic attraction and repulsion.

6. What is a unit charge of static electricity?

7. What is a dielectric? When air is used as the dielectric between two charged bodies, how is the force of attraction or repulsion affected as compared with an equal thickness of paraffin? glass?

8. State the main points of similarity and dissimilarity between electrostatically charged bodies and magnets.

9. A positively charged glass rod is brought near a small sphere. Show by means of a diagram the charged condition of the sphere. Electricity of which sign in this case is on the side near the glass rod?

10. (a) If the charging body (glass rod, Ex. 9) is removed, what will happen to the two charges on the sphere? (b) If before removing the charging body a finger is touched to the sphere or it is otherwise connected for a moment to the earth, what charge will remain on the sphere?

11. In charging a body by induction which should be removed first, the finger or the charging body? Explain.

12. An electrified rod is brought near a pith ball; the ball is first attracted and then repelled. Explain.

13. What is the distinction between conductors and insulators? Name several good conductors; several insulators.

14. State the analogy between water pressure and electric pressure or potential. In what units is water pressure measured? potential?

15. Wherein does an electrometer differ from a simple electroscope? For what is an electrometer used?

16. What is a condenser, and for what is it used? Of what are the conducting sheets of commercial condensers made? What dielectrics are employed?

17. How may the capacity of a condenser be increased?

18. A body is charged, either positively or negatively. Explain why the charge resides upon the surface.

B

19. Make a drawing to illustrate the distribution of a charge (a) on an isolated sphere; (b) on two adjacent spheres; (c) on a body of irregular outline.

20. An "electric whirl" when mounted upon or otherwise in electrical contact with the charged knob of an electric machine will spin around, as illustrated in Fig. 311. Explain.

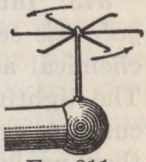


FIG. 311

21. (a) What is lightning? (b) thunder? (c) Which is the cause of the other? (d) What did Franklin prove by the kite experiment?

22. How does a lightning rod protect a building? Wherein is the discharging action of points exemplified by the rod?

23. (a) What is an electric screen, and what practical use is made of it? (b) How can you explain the fact that the modern skyscraper of our cities is so rarely damaged by lightning?

C

24. Explain how to charge a body by means of the electrophorous apparatus. What is the object of touching the fingers to the disc before removing it from the lower plate?

25. What are the main likenesses and differences between the phenomena of electrostatic charges and magnets?

CHAPTER XI

CURRENT ELECTRICITY

The Voltaic Cell

343. Introductory. A *voltaic cell,* sometimes called a *galvanic cell,* is a device for transforming the energy of chemical action into the electrical energy of a current. The lighting, heating, and magnetic effects of electric currents are so commonly met with in everyday life that they no longer excite either comment or wonder. We press a button and the electric light comes on; we turn a switch and the electric flatiron gets hot, or the electric oven is heated; we throw a lever and the electric motor runs, driving machines in our factories, propelling the cars upon our streets, and operating such household devices as the vacuum cleaner, the electric fan, or the sewing machine. Electrical appliances are today everywhere present, and yet it should not be forgotten that they are of comparatively recent origin. As all sciences have developed from the theoretical to the practical, so, too, in this branch of physics, scientists had to study for many years the phenomena associated with electrification, the construction of electric cells, and the nature and effects of electric currents before it was possible to make practical applications of the laws and principles discovered.

While observations on electrostatic phenomena were made and recorded as early as 600 B. C., it was not until 1786 A. D. that it was discovered that an electric current could be produced by chemical action. This important discovery was made by Galvani, professor of anatomy in the University of Bologna, Italy, while he was preparing

frogs' legs for dissecting purposes. The legs and part of the body of a freshly killed frog were suspended by means of a copper wire from the iron frame of a porch in front of his house. The wind caused the moist tissue of the frog to come into contact with the iron frame, thus completing an electric circuit between the frog's legs and the metals, iron and copper. The experimenter noticed that at the instant of contact with the iron frame the frog's legs twitched spasmodically. He concluded that an electric current was generated in some way at a point in the circuit including the iron, copper, and the moist tissue. For a long time after this an electric current which gave rise to convulsive muscular action was known as *galvanic electricity*. We now know that the so-called galvanic electricity does not differ from electricity in general.

That the phenomenon observed by Galvani was caused by a current produced by chemical action was not known for a certainty until 1800, when another Italian scientist named Volta constructed the first electric cell. The term "voltaic cell" was later chosen in honor of Volta.

344. The Simple Voltaic Cell. A voltaic cell such as is frequently used for laboratory experimentation is shown in Fig. 312. Such a cell consists of strips of copper (Cu) and zinc (Zn) partly immersed in a dilute solution of acid in water. The cell is set up as follows: Into a dilute solution of sulphuric acid place a strip of copper and a strip of zinc so that they are a few centimeters apart. The copper becomes positively electrified and the zinc negatively electrified. If now the copper and zinc be connected by means of a wire, a current will flow through the circuit.



Volta

as may be demonstrated by the deflection of the needle of a galvanometer or the ringing of an electric bell when the terminals of these instruments are connected in the external circuit, as at *G*.

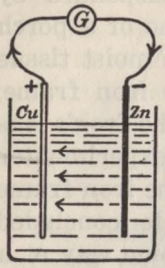


FIG. 312
Simple voltaic cell

This combination of copper, zinc, and acid constitutes what is known as a *simple voltaic cell.* The essential parts of a simple voltaic cell are two metals and a solution, the metals being of such a nature that the solution acts upon one of them more readily than upon the other.

The solution used in a voltaic cell is called the *electrolyte.* The two metals are the *electrodes.* That portion of an electrode which is immersed in the solution is sometimes called the *plate,* and the portion to which the wire is attached, the *pole.* The electrode which becomes positively charged is the *positive electrode*; the one negatively charged, the *negative electrode.* The electrode which is least affected chemically is usually the positive electrode; the one most affected chemically, the negative electrode. For example, in the case of the cell described in the preceding topic, the copper is the positive electrode and the zinc is the negative electrode. The entire path of the current is called the *circuit.* That part outside the liquid is the *external circuit;* the part within the liquid, the *internal circuit.*

When we turn on, or otherwise close a switch so that the current in the system flows uninterruptedly, we then speak of the circuit as a *closed circuit.* A telegraph system operates on a closed circuit, the current flowing practically all of the time. In a door bell system, on the other hand, the current flows only at irregular intervals, as for example, when someone closes the circuit by pressing upon the push button. The door bell circuit is open most

of the time; such a circuit is called an *open circuit.*

An *electric battery* consists of a single cell or a number of cells joined together.

345. Charges on the Electrodes. When electrodes, such as copper and zinc, are immersed in an electrolyte, the copper becomes positively charged and the zinc negatively charged by the chemical action which takes place in the cell. The signs of the charges on the respective electrodes of a cell may easily be determined by means of a sensitive electroscope. Use is made of an electroscope which has upon the upper end of the conductor carrying the gold leaves a circular metal disc, 6 or 8 centimeters in diameter, Fig. 313. A second and similar disc is provided with a glass insulating handle. Place one disc upon the other, separating them from each other by means of a sheet of paper, or better, a thin sheet of mica. The two discs so placed constitute a condenser (Art. 334). By means of wires, connect one electrode to the lower disc and the other eletrode to the upper disc, as shown. A charge of one sign flows from the cell to the lower disc, and a charge of opposite sign appears upon the upper disc. Now remove the wires from the discs, and then remove the upper disc. The charge which was "bound" upon the lower disc by the opposite charge upon the upper disc now being free, distributes itself over the conducting systems of the electroscope, thus causing the leaves to diverge. To test the charge upon the leaves,

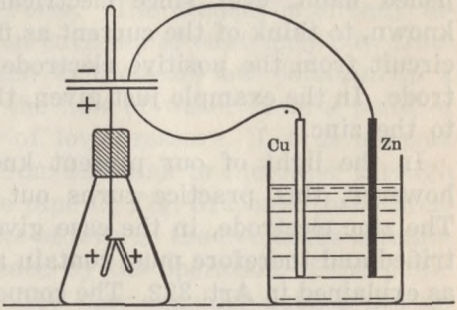


FIG. 313.—Positive charge on copper; negative charge on zinc

touch the top of the electroscope with a proof plane carrying a known charge (a positive charge from glass, for example). If the leaves fly farther apart we know that the charge imparted to them from the electrode of the cell is positive; if, on the other hand, the leaves fall nearer together then we know that the charge upon the electrode is negative.

By means of a test such as just described we learn that the charge on the copper is positive and that on the zinc is negative.

346. Direction of the Current. It has been the established habit, ever since electrical currents have been known, to think of the current as flowing in the external circuit from the positive electrode to the negative electrode. In the example just given, this is from the copper to the zinc.

In the light of our present knowledge of electrons, however, this practice turns out to be embarrassing. The zinc electrode, in the case given, is negatively electrified and therefore must contain an excess of electrons, as explained in Art. 322. The copper electrode, positively electrified, has a deficiency of electrons. Now the electric current is, according to present notions, a stream of electrons. It therefore will flow from the region with an excess of electrons and negatively electrified to the region with a deficiency of electrons and positively electrified. The electric current flows through the metal conductor, but exactly in the reverse direction to that which we are accustomed to give.

Meanwhile, there has been written, printed, read, and used an immense number of books on all aspects of electric currents, using the customary way of describing the direction of currents. In all the practical books on electricity, including electricians' manuals, handbooks on radio, etc., the currents are described as passing through

the conductor from the positive electrode to the negative one. The attempt to change our way of describing the current now would simply produce hopeless confusion. In this text, therefore, we shall continue to follow the established custom, and will speak of currents as passing in the external circuit from the positive to the negative electrode. The student should bear in mind, however, that the stream of electrons actually flows in a direction exactly opposite to that of the current as described. The student is cautioned not to worry about this, but simply to remember the fact.

347. Electromotive Force. The student is already familiar with the idea that currents of electricity flow from points of high potential to those of low potential in a manner analogous to the flow of water from a point of high pressure to one of low pressure. Let us suppose that the water flows from one tank to the other through the pipe *C*, Fig. 314, and that difference of level in the two tanks is maintained by the operation of the pump *P*. We may consider such a system as somewhat analogous to that of an electric cell, the difference in pressure between *A* and *B* corresponding to the difference of potential between the electrodes of the cell. In the case of the water the difference of level is maintained by the expenditure of energy through the agency of the pump; *in the electric cell the difference of potential is maintained by the expenditure of energy due to chemical action.*

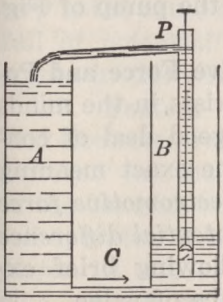


FIG. 314
Analogy between hydrostatic pressure and e.m.f.

The *electromotive force* (e.m.f.) of an electric generator (battery or dynamo) is its capacity for generating electric pressure or potential. Electromotive force

is that which produces a difference of potential in a circuit, and therefore it is the cause of the flow of electric currents. The greater the e.m.f. of a battery the greater the current it will produce in a conductor of given electrical resistance. The unit of e.m.f. is the *volt*, the unit of current strength the *ampere*, and the unit of resistance the *ohm*. The precise definitions of these units will be given later.

With reference to the term *electromotive force*, it is important to make note of two things: First, while we speak of electric pressure as being *analogous* to water pressure, it must be borne in mind that *electromotive force is not a force*, nor is it measured in units of force; *e.m.f. is measured in units of work (ergs) per unit quantity of electricity conveyed around the circuit.* And in the second place, a voltaic cell does not generate electricity; it generates *electromotive force.* The cell does not generate electricity any more than the pump of Fig. 314 generates water.

348. Distinction between Electromotive Force and Potential Difference. Since there often exists in the minds of students a good deal of confusion as to the exact meaning of the terms *electromotive force* (e.m.f.) and *potential difference* (p.d.), the following brief explanation may be of value.

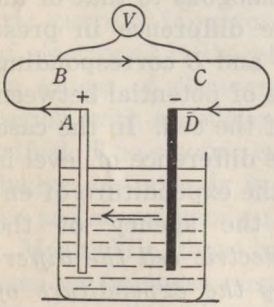


FIG. 315
Electromotive force and difference of potential

Electromotive force and potential difference are both measured in *volts*. Let us consider a circuit, Fig. 315, which includes an electric cell with its internal circuit and the external conductor *ABCD*. The circuit is closed and a current flows around the circuit. If a volt-

meter V be connected to the points B and C it will measure the fall of potential (p.d.) in volts between these points; in a like manner we may determine the p.d. between AB or CD or any other two points we choose in the conductor AD.

Now the e.m.f. of the cell is equal to the sum of all the potential differences around the *entire circuit*, including both that of the external circuit and the internal circuit; that is, in this case e.m.f. = p.d. from A to B + p.d. from B to C + p.d. from C to D + p.d. of the internal circuit. The p.d between A and D over the external circuit is called the *terminal potential* of the battery; it is always somewhat less than the e.m.f. of the battery. The exact e.m.f. may be determined by means of an electrometer (Art. 333) which is calibrated to give readings in volts. Electromotive force then may be thought of as a *cause*; potential difference as an *effect*. Thus the e.m.f. of a battery is the cause of the flow of the current; the p.d. is the fall of potential (measured in volts) between any two points in the circuit.

349. Factors that Determine the E.M.F. of a Cell. The electromotive force of a cell is determined by three factors: (a) the kinds of electrodes used, (b) the nature of the electrolyte, and (c) the temperature at which the cell operates. *The e.m.f. is independent of the size of the plates.* A cell of given materials, such for example as zinc, copper, and dilute sulphuric acid, no larger than a lady's thimble will furnish as high an e.m.f. as will a similar cell having a volume capacity of several gallons.

While the e.m.f. of a cell is independent of its size, nevertheless a large cell is, in general, more desirable than a small one, because the larger the cell the less the resistance, and hence the greater the current, and also the larger the cell the longer will its materials last to furnish a current.

350. Voltaic Cells with Different Electrolytes. It must not be supposed that in order to make a voltaic cell it is necessary to use a dilute solution of sulphuric acid, as the electrolyte. Any solution, be it acid, base, or salt, which will act chemically upon the one electrode more readily than upon the other, will serve as an electrolyte. This may be shown by the following experiment. Use electrodes of copper and zinc, connecting the poles to a galvanometer. Dip these electrodes successively into the following solutions, rinsing the electrodes in water after each test: (a) dilute sulphuric acid, (b) dilute hydrochloric acid, (c) dilute acetic acid (vinegar), (d) solution of sodium chloride (NaCl). In each case the galvanometer is deflected, showing that a current is set up by the cell. If the electrodes be thrust, for example, into an apple, Fig. 316, a current will also flow, due to the action of the juices of the apple upon the metals. In fact, almost any vegetable will act chemically upon the zinc sufficiently to set up a current.

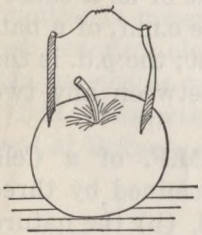


FIG. 316
An apple cell

351. Voltaic Cells with Different Electrodes. In this experiment it is desired to show that different substances may be used as electrodes. Take as the electrolyte a dilute solution of sulphuric acid. Let us consider a series of electrodes, such as zinc (Zn), iron (Fe), copper (Cu), silver (Ag), platinum (Pt), and carbon (C). First, use as electrodes copper and zinc. The copper, as we have seen, is positive to the zinc, causing the pointer of the galvanometer to be deflected in a given direction. Second, use as electrodes copper and carbon. The current now flows through the galvanometer in the opposite direction; that is, from the carbon to the copper, as shown by the opposite deflection of the pointer. The

copper is in this case the negative electrode. By similar experiments it may be shown that copper is positive to zinc and iron, but is negative to silver, platinum, and also, as has been shown, to carbon.

In the following list any element, copper for example, is positive to the elements on the left of it and is negative to elements on the right:

— Zn Fe Cu Ag Pt C +

Thus iron is positive to zinc and negative to copper, while copper is positive to iron, but negative to silver. The farther any two elements in the list are removed from each other, the greater is the terminal potential difference of the cell. In most commercial cells, therefore, zinc and carbon are used as electrodes.

352. Local Action. If a strip of commercial zinc be placed in a dilute solution of sulphuric acid, chemical action at once takes place. The zinc is dissolved, forming zinc sulphate and liberating hydrogen. The reaction may be written $Zn + H_2SO_4 = ZnSO_4 + H_2$, which represents one atom of zinc (Zn) uniting with one molecule of sulphuric acid (H_2SO_4) to form one molecule of zinc sulphate ($ZnSO_4$) and two atoms of gaseous hydrogen (H_2). This chemical action is called *local action*; it is due to impurities in the zinc in the form of tiny particles of carbon, iron, etc. Each particle constitutes, with the zinc and acid, a voltaic cell, Fig. 317. If the zinc were perfectly pure there would be no local action.

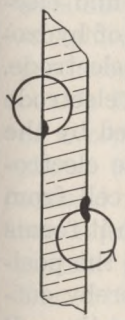


FIG. 317
Local action

Local action is detrimental because it represents a useless waste of zinc, giving rise to chemical action, but furnishing no useful current to the circuit. It may be prevented by coating the zinc plate with mercury; that is, *amalgamating* the zinc.

To demonstrate the effect of amalgamating the zinc plate, take a strip of commercial zinc and after cleaning it by immersing it for a few moments in dilute sulphuric acid, rub mercury over its surface until the latter becomes bright and mirror like. The impurities in the zinc are now covered up by the mercury, which not only spreads over every part of the surface of the plate but also penetrates to a considerable depth. If the amalgamated plate thus formed be now placed in the acid, no chemical action will occur, so long as the zinc is not a part of the electrical circuit of the cell. If, however, the amalgamated zinc plate be connected to the copper electrode by means of a wire, chemical action will at once take place between the acid and the amalgamated zinc, all the energy being now used to furnish a current. Thus it appears that amalgamating the zinc plate prevents local action, and, at the same time does not hinder in any way the zinc from acting as the negative electrode when connected with the positive electrode by means of a wire.

353. Polarization. If we close the circuit of a simple voltaic cell consisting of dilute sulphuric acid and electrodes of copper and amalgamated zinc, bubbles of hydrogen will appear in large numbers at the copper electrode. This appearance of hydrogen on the positive electrode does not mean that the copper is being dissolved by the acid; it does mean that the hydrogen from the electrolyte is transferred during the operation of the cell from the solution to the positive electrode, upon which it forms a gaseous layer. This collection of hydrogen on the positive electrode reduces the e.m.f. of the cell, thereby cutting down the current. When in this condition the cell is said to be polarized.

Polarization is the reduction of the e.m.f. of a cell, due usually to the collection of hydrogen on the positive electrode.

Polarization is detrimental because it reduces the e.m.f. of the cell, and also increases its resistance. For both these reasons polarization always tends to reduce the current which a cell will furnish. It may be remedied by adding to the electrolyte an oxidizing agent, that is, a substance which will furnish oxygen to unite chemically with the hydrogen on the electrode and thus form water, as represented by the chemical equation

$$2\,H + O = H_2O$$

KINDS OF CELLS

354. Introductory. There are many different types of cells (batteries) depending on (a) their form, (b) the character of the electrodes, and (c) the nature of the electrolyte. In this text only a few of the more common types will be considered, namely, the Daniell and gravity cells, the Leclanché cell, and the dry cell.

So far as the consumption of energy is concerned as compared with the useful work performed, the voltaic cell is a rather wasteful engine. Wet cells, that is, voltaic cells having as an electrolyte a solution of an acid, base, or salt, are gradually being driven out of use as current producers by the more convenient "dry cell."

355. Daniell and Gravity Cells. A *Daniell cell* consists of a copper plate immersed in a solution of copper sulphate ($CuSO_4$) and a zinc plate immersed in zinc sulphate ($ZnSO_4$) the two liquids being kept apart by an unglazed, porous cup. The *gravity cell*, Fig. 318, is in principle a Daniell cell in which the two solutions ($CuSO_4$ and $ZnSO_4$) are separated by gravity, the copper sulphate solution being the heavier of the two.

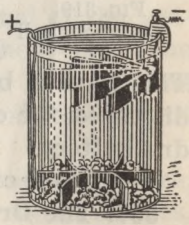


FIG. 318
Gravity cell

In both cells copper is the positive

electrode and zinc the negative. Both cells also contain two different solutions, namely, a saturated solution of copper sulphate in contact with the copper plate, and a dilute solution of zinc sulphate in contact with the zinc plate. Batteries of this type are called two-fluid batteries. When the cell is in operation zinc goes into solution at the negative (Zn) electrode and copper comes out of solution at the positive (Cu) electrode. Thus as times goes on the zinc electrode grows lighter and the copper electrode heavier. Under ordinary conditions no hydrogen is liberated on the copper plate and consequently no polarization occurs.

Cells of this type are therefore useful for *closed circuit* work, that is, where a current of fairly constant e.m.f. is desired.

The e.m.f. of Daniell and gravity cells is about 1 volt.

356. The Leclanché Cell. The electrolyte of the Leclanché cell, Fig. 319, is a solution of ammonium chloride. A rod of carbon serves as the positive electrode; a rod of zinc, as the negative electrode. The positive carbon electrode is placed in a porous cup which is filled with a mixture of graphite and manganese dioxide. The function of the manganese dioxide is to serve as the depolarizing agent.

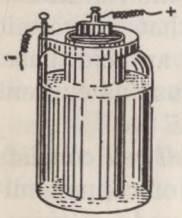


FIG. 319
Leclanché cell

This cell is used for *open circuit work,* such as the ringing of door bells. This type of battery, however, has been almost entirely displaced for open circuit work by the more convenient dry battery.

The e.m.f. of the Leclanché cell is about 1.4 volts.

357. The Dry Cell. The so-called dry cell, Fig. 320, is a modified form of the Leclanché cell. In reality it is not a dry cell at all, since it contains a paste in which there

is always a certain amount of water. The outer case of the cell consists of a cylindrical zinc can, which serves as the negative electrode; the positive electrode is a carbon rod which is placed in the center of the cell. Inside the zinc container there is a sheath of wood pulp which is saturated with the electrolyte. The carbon rod is surrounded by a mixture of granular carbon and manganese dioxide, the object of which is to serve as a depolarizer for the carbon electrode.

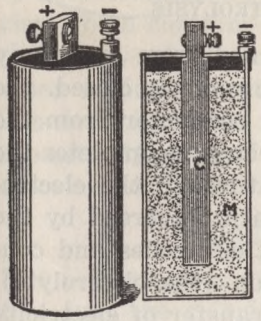


FIG. 320.—Dry cell

The electrolyte is a moist paste containing ammonium chloride (NH_4Cl), commonly called sal ammoniac, and a certain amount of zinc chloride ($ZnCl_2$). The zinc chloride is added to reduce the corrosive action of the ammonium chloride when the cell is idle. The cell is hermetically sealed to prevent evaporation.

The e.m.f. of the dry cell is about 1.5 volts

The standard size of dry cells is, diameter 2.5 inches, height 6 inches. Smaller cells are made for use in flashlights. A good cell of standard size should give a discharge of about 25 ampere hours; that is, 1 ampere for 25 hours, or 0.5 ampere for 50 hours. Dealers use a small ammeter for testing their cells, one terminal of the testing instrument being connected with the — pole, the other being pressed against the + pole for a short time until a reading can be taken. A new cell should test at least 15 amperes.

Dry batteries are well suited for intermittent service, such as door bells, flashlights, telephones, signal devices, and the ignition circuits of certain types of gas engines. An enormous number of dry batteries are used, about 50

million being manufactured in this country each year.

Dissociation and Electrolysis

358. Introductory. We have learned from our study of the voltaic cell that when the circuit is closed, the current which flows in the external conductor from the positive electrode to the negative electrode completes the circuit from electrode to electrode through the electrolyte. That is to say, part of the path traversed by the current is a solid conductor (metal electrodes and connecting wire) and part is a solution (the electrolyte). Now the mechanism by which the transfer of electricity takes place in the wire is quite different from that which is associated with the transfer of the electricity through the liquid of the cell. In the wire, transfer of electricity (current) is due to a stream of electrons which are more or less impeded in their motion by the intervening molecules of the metal. In the solution (electrolyte), on the other hand, the electrical transfer is due to the bodily displacement of charged particles; in other words, the electric charges are carried bodily from one electrode to the other through the solution by particles called ions.

It thus appears that as a conductor of electricity an electrolyte, as compared with a metallic conductor, possesses certain distinctive and important features. In order to learn what goes on in an electric cell during the passage of a current, especially as relating to electrolysis and the storage battery, it will be necessary for us to learn something about ions and their properties.

359. Ionization. Whenever acids, salts, and certain other chemical compounds are put into water two things happen: First, they dissolve in the water, that is, they go into solution; second, they dissociate into positively and negatively charged ions. *Dissociation* is the breaking up of the molecules of a substance in solution into

ions. An *ion* is an atom or group of atoms carrying one or more charges of electricity. An *electrolyte* is a solution which contains ions.

By way of illustration let us consider what happens when sodium chloride (NaCl) or common table salt is dissolved in water. The molecule NaCl breaks up into a positively charged sodium ion and a negatively charged chlorine ion. The reaction may be represented as follows:

$$\text{NaCl in water} = \overset{+}{\text{Na}} + \overset{-}{\text{Cl}}.$$

The sodium atom with the + sign above it is the positive ion; the chlorine atom with the — sign, the negative ion. The negative ion in this case has an excess of one electron; the positive ion has a deficiency of one electron, hence its positive charge. A careful distinction here should be made between *ions* and *electrons*. An ion is an atom or group of atoms with one or more charges of electricity. The charge on the ion is due to the excess or deficiency of electrons.

Hydrochloric acid (HCl) and sulphuric acid (H_2SO_4) are examples of acids; silver nitrate ($AgNO_3$) and copper sulphate ($CuSO_4$) are examples of salts. In each case the symbols appearing within the parentheses represent the chemical composition of the respective molecules. The dissociation of these acids and salts are given in the following table:

$$\begin{aligned}
\text{HCl} \quad \text{in water} &= \overset{+}{\text{H}} + \overset{-}{\text{Cl}} \\
\text{H}_2\text{SO}_4 \quad \text{“} &= \overset{+}{\text{H}} + \overset{+}{\text{H}} + \overset{--}{\text{SO}_4} \\
\text{AgNO}_3 \quad \text{“} &= \overset{+}{\text{Ag}} + \overset{-}{\text{NO}_3} \\
\text{CuSO}_4 \quad \text{“} &= \overset{++}{\text{Cu}} + \overset{--}{\text{SO}_4}
\end{aligned}$$

When a molecule dissociates into ions there is not necessarily an equal number of positive and negative ions

formed, *but there is always an equal number of* + and — *charges.* Consider H_2SO_4, as given in equational form above. When a molecule of sulphuric acid (H_2SO_4) ionizes, there are formed three ions, two H ions each carrying a + charge and one SO_4 ion carrying two — charges. When copper sulphate ($CuSO_4$) ionizes there are formed only two ions, the Cu ion carrying two + charges and the SO_4 ion carrying two — charges.

360. Theory of the Voltaic Cell. Having learned something of ionization, we are now prepared to offer an

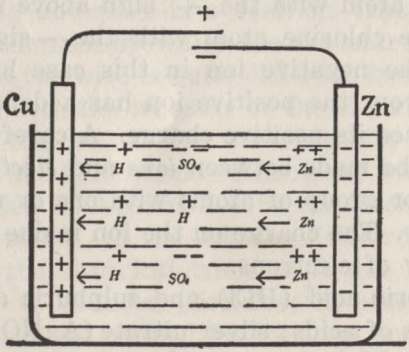


FIG. 321.—Electrochemical action in a voltaic cell.

explanation as to why the zinc electrode in a battery becomes negatively charged and the copper or carbon electrode positively charged. Consider a cell having copper-zinc electrodes immersed in an electrolyte consisting of a dilute solution of sulphuric acid, Fig. 321. The electrolyte contains positively charged H ions and negatively charged SO_4 ions.

Now when the zinc plate is put into the electrolyte, positively charged Zn ions pass from the zinc electrode to the solution. This leaves an excess of electrons (— charges) in the zinc rod which accounts for its — charge. If the battery circuit is open Zn ions continue to leave the

zinc plate until the pulling force (attraction) of the negatively charged zinc plate for the positively charged Zn ions is just sufficient to balance the tendency of the zinc to go into solution. Some of the $-SO_4$ ions of the acid unite with the + Zn ions forming zinc sulphate ($ZnSO_4$).

Let us inquire now how the copper plate gets its + charge. There are in the electrolyte large numbers of positive H ions. When the Zn ions go into solution they give the liquid around the zinc electrode a strong positive charge, which exerts a repellent force on the positive H ions, driving them to the copper electrode. The passage of H ions to the copper continues until the pushing effect (repulsion) of the now positively charged copper plate is just sufficient to counterbalance the tendency of the positive H ions to collect upon it. The copper plate, then, gets its + charge from the positively charged H ions.

When the battery is on closed circuit, that is, when a current is flowing, the conditions of equilibrium at the zinc and copper plates, as described above, are broken, and consequently so long as the current continues to flow Zn ions go into solution at the negative electrode and H ions come out at the positive electrode.

361. The Electrolytic Cell. Suppose that we have a cell such as shown in Fig. 322, consisting of two similar electrodes and an electrolyte. Let the electrodes be of platinum and the electrolyte be a dilute solution of hydrochloric acid (HCl). The object of using platinum electrodes in that this metal will not dissolve in any of the commonly used acids. The HCl dissociates into positively charged H ions and negatively charged Cl ions. Now if a current be passed through this cell in a direction indicated

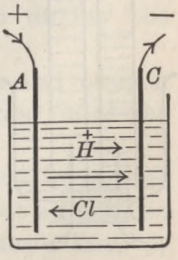


FIG. 322
Electrolytic cell

by the arrows at *A* and *C*, the + ions go *with* the current and the — ions go *against* the current. The electrode toward which the + ions go is called the *cathode*; the other electrode, the *anode*. In the electrolyte cell shown in Fig. 322, *C* is the cathode, *A* the anode. It is characteristic of H ions and metallic ions to go with the current; negative ions always go against the current.

Such a cell as here described is called an *electrolytic cell*. It differs from a voltaic cell in this respect: A voltaic cell is designed to *furnish* a current; an electrolytic cell is designed to *receive* a current from some outside source, thus causing a separation of the ions at the electrodes. The — ions appear at the anode and the + ions at the cathode.

Electrolysis is the process by which the ions (+ and —) of a solution are separated by means of a current.

362. The Electrolysis of Water. The electrolysis of water furnishes a most instructive illustration of the electrolytic effects of a current. In this experiment a cell of the type shown in Fig. 323 may be used. The electrodes are of platinum. The electrolyte is a dilute solution of sulphuric acid (H_2SO_4) in water. When the current passes through such a cell, the H ions appear at the cathode, give up their positive charges, and are liberated in the form of gaseous hydrogen. Oxygen, O, is liberated at the anode. The relation of the volume of O to the volume of H is as 1 to 2; that is, for every cubic centimeter of O liberated at the anode there are two cubic centimeters of H liberated at the cathode.

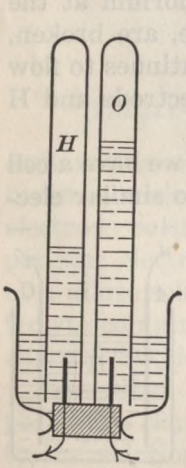


FIG. 323.—Electrolysis of water

The chemical reactions which take place in the electrolysis of water con-

taining sulphuric acid are quite complex, the nature of the reaction depending upon the e.m.f. applied. A full discussion of this subject is beyond the limits of an elementary text, it being sufficient to say that the object of adding sulphuric acid to the water is to increase the concentration of the ions and thus to make the solution more conducting.

Decomposition of the water molecules occurs in this electrolytic process largely as a secondary reaction, as follows: The acid on being added to the water is dissociated into positive ions, and negative ions, thus

$$H_2SO_4 \text{ in water} = \overset{+}{H} + \overset{+}{H} + \overline{\overline{SO_4}}$$

The H ions are discharged at the cathode, forming gaseous hydrogen; the SO_4 ions appear at the anode, giving up their negative charges and liberating gaseous oxygen, probably after this manner:

$$SO_4 + H_2O = H_2SO_4 + O$$

This reaction means that the SO_4 unites with a molecule of water, H_2O, forming one molecule of H_2SO_4 and liberating one atom of O at the anode. The sulphuric acid thus formed again dissociates forming more H and SO_4 ions. And so, the process continues. It will be noted that there is just as much sulphuric acid in the solution when we get through as when we started. The water molecules, however, are constantly being broken up during the process.

The electrolysis of water into hydrogen and oxygen for commercial purposes is extensively carried on at the present time.

363. The Laws of Electrolysis. The laws governing the deposition of ions by electrolysis were first announced by Faraday and are now known as *Faraday's laws of electrolysis*. They are two in number and may be stated as follows:

I. *The mass of a substance deposited upon the electrodes of an electrolytic cell is proportional to the strength of the current and to the time which it flows.* This means that if a given current flowing through an electrolytic cell for 1 hour deposits 1 gram of copper, then twice as much current flowing for 2 hours will deposit $2 \times 2 = 4$ grams of copper.

II. *If the same quantity of electricity be passed through a number of electrolytic cells, each containing a different electrolyte, the mass of each substance deposited is proportional to its chemical equivalent.* The *chemical equivalent* of an element is its atomic weight divided by its valence. The relation of atomic weights to chemical equivalents for hyrogen and oxygen may be shown as follows:

Element	Atomic Weight	Valence	Chemical Equivalent
H	1	1	1
O	16	2	8

For example, when, in an electrolytic cell such as is shown in Fig. 323, 1 gram of hydrogen is liberated at the cathode, 8 grams of oxygen are liberated at the anode. The student is cautioned not to confuse the relative volumes of these elements liberated (Art. 362) with their respective masses. For the passage of a given quantity of electricity, the respective *volumes* are H:O = 2:1, while their *masses* are H:O = 1:8.

364. Practical Applications of Electrolysis. *Electroplating* is the process of coating one metal with another by electrolytic action. An electroplating tank is nothing more than a large electrolytic cell in which the electrolyte is a solution containing the ions of the metal required and in which the cathode is the object to be plated. In order that the concentration of the metallic

ions in the electrolyte remain constant it is necessary to use as the anode a bar or plate of pure metal similar to that which goes out of solution at the cathode. For example, if we are silverplating we shall have to use as the anode a piece of pure silver; for nickelplating, an anode of nickel, and so on.

Silverplating. The silverplating of tableware, such as spoons, knives, forks, etc., Fig. 324, is a familiar illustration of the electrolysis of a silver salt. For such work a solution of silver cyanide is used as the electrolyte, since solutions of this salt give a smooth and compact deposit. The electrolysis of silver nitrate, $AgNO_3$, also furnishes a standard method of measuring a current of electricity, as will be explained later in Art. 371.

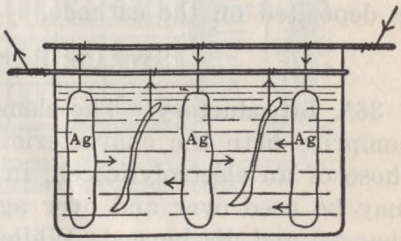


FIG. 324.—Silverplating cell

Nickel plating. The plating of bicycle and automobile parts and of household fixtures to protect articles made of brass, iron, or steel is very extensively carried on. The electrolyte is a solution of nickel ammonium sulphate.

Copper plating and electrotyping. In copper plating and electrotyping a bath of copper sulphate is employed. A piece of copper serves as the anode. The object to be plated is suspended as the cathode. In electrotyping an impression of the type, after being set up, is made in wax, the surface of which is coated with finely powdered graphite to render it conducting. The wax mold thus prepared is now put into the bath as the cathode, and a suitable deposit of copper made upon it.

Electrolytic refining of copper. As obtained from its

ores, copper contains a number of impurities which make it objectionable for certain kinds of electrical service. To remove these impurities the electrolytic process of refining is largely employed. A large piece of impure copper is suspended as the anode in a bath of copper sulphate; a small piece of pure copper is used as the cathode. A current of proper strength is passed through the cell. The anode gradually dissolves and pure copper is deposited on the cathode.

Storage Batteries

365. Introductory. The elements of a storage battery comprise both the characteristics of a voltaic cell and those of an electrolytic cell, in that the storage battery may be used over and over again by being alternately charged and discharged. While being charged it acts as an electrolytic cell, and on discharge it acts as a voltaic cell.

Considering its extended use in automobile lighting and starting systems, as an auxiliary in electric lighting plants, and in technical laboratory service, we have in the storage battery one of the most important appliances of modern science, millions of dollars being invested today in the manufacture and installation of battery systems of this sort. Among the many practical uses of the storage battery, in addition to those just mentioned, is its adaptation to meet the requirements of modern naval practice in the development of the submarine. The submarine can live in warfare only by stealth and so in order to steal through the water without awakening a suspicion of its presence, it must be equipped with propelling power which is practically noiseless and which leaves no visible trace of its operation. The storage battery and the motor it energizes are almost ideally suited to the requirements of the submarine, for they

are silent and practically heatless, leaving no trace of bubbles to betray the whereabouts of the submerged vessel. Some idea of the size of the battery required in

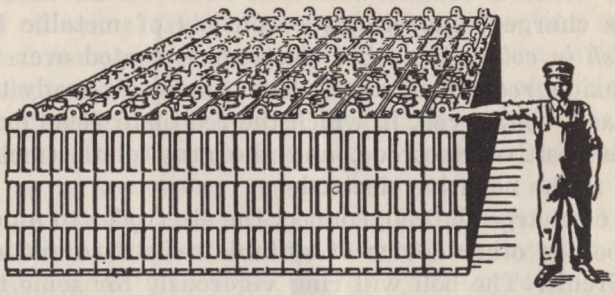


FIG. 325.—One of the storage battery units of a modern submarine, consisting of 60 giant cells

our large submarines may be had from the accompanying sketch, Fig. 325.

366. The Lead Storage Battery. On charge, a storage battery is an electrolytic cell; on discharge it is of the nature of a voltaic cell. On charge it receives a current; on discharge it furnishes a current.

The essential features of the operation of a storage cell may be illustrated by means of a simple apparatus, Fig. 326, which consists of two pieces of sheet lead immersed in a dilute solution of sulphuric acid.

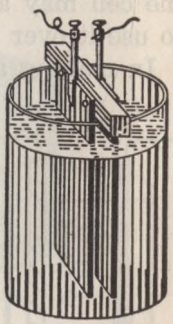


FIG. 326
Simple storage cell

The first step is to charge the cell. This is done by passing a current through the cell for a short time. During the process of being charged, the cell acts as an electrolytic cell, hydrogen being liberated at the cathode and oxygen at the anode. The hydrogen unites with any oxygen that may be associated with the

cathode, thus reducing the latter to metallic lead, of a dull gray color. The oxygen, on the other hand, unites with the lead of the anode, oxidizing its surface to lead oxide, which is reddish brown in color. Thus when the cell is charged, one electrode consists of metallic lead, *grayish in color;* the other electrode is coated over with lead oxide, *reddish in color.* The cell is now ready to be used as a voltaic cell, in which the red plate (lead oxide) is the positive electrode, and the gray plate (metallic lead) is the negative electrode.

To discharge the cell, connect the electrodes to an electric bell or other device of suitable resistance and close the circuit. The bell will ring vigorously for some time, that is, until the cell is discharged.

During discharge the current in the cell flows in the reverse direction from that of charge. This means that the hydrogen ions now go to the red oxidized plate, uniting with the oxygen to form water, thus reducing the surface of the plate to metallic lead. After discharge, the cell may again be charged, thus making it possible to use it over and over again.

In making a study of this simple storage battery the student should note several things, as follows: (a) When the cell is being charged, the charging current passes through the cell in one direction (from left to right, say), and on discharge the current furnished by the cell flows in the opposite direction, that is, the electrode which is positive on charge becomes neg-

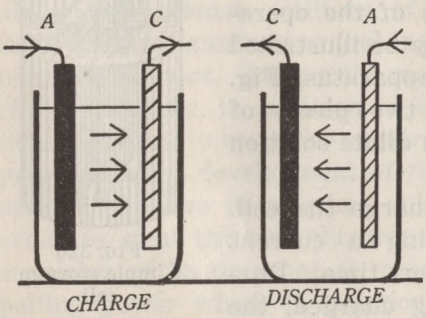


FIG. 326*a*.—Storage battery on charge and discharge

ative on discharge. (b) When the cell is charged the red plate is the positive electrode and the gray plate is the negative electrode. (c) The hydrogen and oxygen liberated at the electrodes are furnished by the molecules of water (H_2O), the sulphuric acid acting merely as an intermediary agent, as explained in the article on the electrolysis of water (Art. 362). The sulphuric acid concentration is diminished in discharge, but it is completely restored on charge. (d) A storage battery does not, as is sometimes supposed, store up electricity; a storage battery is merely a device for storing up energy. On charging the battery the energy of an electric current is converted into potential energy by the reduction of the cathode and the oxidation of the anode.

367. Characteristics of the Lead Storage Battery. The electrodes of a commercial lead storage battery are shown in Fig. 327. The electrodes consist of two series

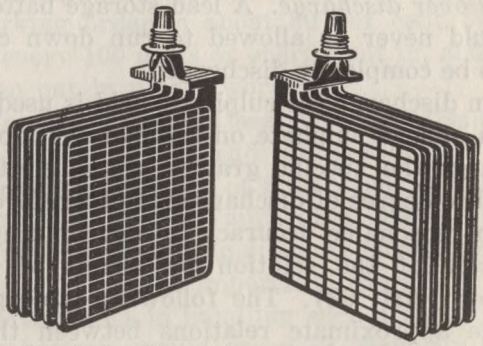


FIG. 327.—Electrodes of a lead storage battery

of lead plates, with which a small quantity of the metal antimony (Sb) has been united as an alloy. The object of adding the antimony to the lead is to give stiffness and strength to the latter. The electrode which is positive on discharge is thickly coated with lead peroxide;

indeed, in some types of the lead cell, lead peroxide is forced under high pressure into holes or grids which have previously been punched in the plate.

When the battery is charged, the chemical condition of the electrodes may be represented as follows:

Positive plate (red) = lead peroxide (PbO_2)
Negative plate (gray) = metallic lead (Pb)

On discharge lead sulphate ($PbSO_4$) is formed on both plates. The formation of lead sulphate is a necessary condition for the effective operation of a storage cell. The lead sulphate crystals are broken up chemically when the cell is recharged. Sometimes, however, this decomposition of the lead sulphate does not take place completely, in which case the cell is said to be *sulphated.* This formation of permanent lead sulphate crystals on the plate proves very troublesome and seriously reduces the efficiency of the battery. *Permanent sulphating is usually caused by over discharge.* A lead storage battery, therefore, should never be allowed to run down completely, that is, to be completely discharged.

Since on discharge the sulphuric acid is used up in the formation of lead sulphate on the plates, it follows that on discharge the specific gravity of the electrolyte decreases. The degree of discharge may therefore be measured with reasonable accuracy by determining the specific gravity of the solution by means of a syringe hydrometer (Art. 62). The following table gives at a glance the approximate relations between the specific gravity and the degree of charge and discharge:

specific gravity about 1.29 = full charge
specific gravity about 1.25 = half charge
specific gravity below 1.20 = discharged

In this connection it should be noted that while the sulphuric acid content of the electrolyte is depleted due to

the formation of lead sulphate during discharge, the acid concentration is again restored on charge. This means that under ordinary conditions no acid has to be added to the battery. Water should be added from time to time, however, to replace that lost by evaporation. The chemical reactions for charge and discharge may be represented thus:

$$\text{Charge, } 2\,PbSO_4 + 2\,H_2O = PbO_2 + Pb + 2\,H_2SO_4$$
$$\text{Discharge, } PbO_2 + Pb + 2\,H_2SO_4 = 2\,PbSO_4 + 2\,H_2O$$

This means that on charge lead sulphate ($2PbSO_4$) breaks down forming lead peroxide (PbO_2) at one electrode and spongy lead (Pb) at the other, sulphuric acid (H_2SO_4) also being formed, and on discharge lead sulphate is formed and the sulphuric acid concentration is diminished.

The e.m.f. of the lead storage battery when fully charged is high, namely, 2.1 volts. Its efficiency when in good working order is about 80 per cent. This means that for every 100 units of energy put in, about 80 units are gotten out in useful work.

368. Precautions to be Observed in the Use of the Lead Storage Battery. The lead storage battery is a very efficient and useful piece of apparatus; it is also a relatively expensive piece of apparatus. Its life and usefulness may be seriously impaired by careless or ignorant treatment. It is very important, therefore, to observe the following rules in handling such a battery.

1. The battery should never be short circuited.

2. The battery should not be overcharged. Overcharging causes overheating, which in turn causes the plates to bend or "buckle," thus crowding the lead peroxide out of the grids or holes, in the peroxide plate. On charging, when the cells begin to "gas" freely the current should be cut off; when a storage cell begins to give off

hydrogen gas freely it may in general be considered to be fully charged.

3. The battery should not be over-discharged, that is, discharged below an e.m.f. of about 1.8 volts, or below a sp.g. of 1.25. Over-discharge causes the plate to become permanently sulphated.

4. Only pure distilled water should be used in making up the sulphuric acid solution as the electrolyte. Metallic impurities in the water used in a battery of this sort are fatal to the life of the cells.

5. Pure water should be added to the battery from time to time in sufficient quantities to keep the plates completely submerged. The electrolyte should just cover the top of the plates.

6. Storage batteries when not in use should be frequently charged.

369. The Edison Storage Battery. The nickel-iron storage cell was invented by Thomas Edison, hence the name "Edison storage cell." The positive electrode consists of a nickel plate, and the negative electrode consists of an iron plate, both being immersed in a solution of caustic potash (KOH) or caustic soda (NaOH). This electrolyte contains a solution of 21 per cent of caustic in pure distilled water. The chemical reactions of the Edison cell are not so well known as are those which occur in the lead-sulphuric acid cell. On charge the nickel plate is oxydized to nickel oxide, (NiO_2). On discharge the nickel oxide is reduced, and the iron electrode is oxidized to iron oxide, (FeO). It is said that the Edison cell is not appreciably injured by short circuiting. One of the objects which the inventor had in mind when his cell was devised was to make it as nearly "fool proof" as possible.

The e.m.f. of the Edison battery is about 1.2 volts. Its efficiency is somewhat below that of the lead cell. For

equal current capacities, however, the Edison battery weighs only about half as much as a lead battery.

Units of Electrical Quantities

370. Introductory. The four most important units employed in measuring electrical quantities are (a) the ampere, the unit of current strength; (b) the ohm, the unit of resistance; (c) the volt, the unit of potential difference and e.m.f., and (d) the coulomb, the unit of quantity.

Nearly all the units employed in the measurement of electrical quantities have been named in honor of some celebrated physicist. The ampere, for example, was so named in honor André Marie Ampère, a noted French physicist; the ohm after Ohm, a German physicist. Likewise the volt and coulomb were named in honor of Volta and Coulomb, respectively, the former an Italian and the latter a Frenchman.



Ampère

371. The Ampere and the Coulomb. The ampere is the unit of current strength and the coulomb the unit of quantity. A standard and fundamental method of measuring a current of electricity is furnished by the electrolysis of silver nitrate, $AgNO_3$. The electrolytic cell employed is called a silver voltameter, or better, a *silver coulometer*, since it measures the quantity of electricity that passes through it. One form of the silver coulometer is shown in Fig.

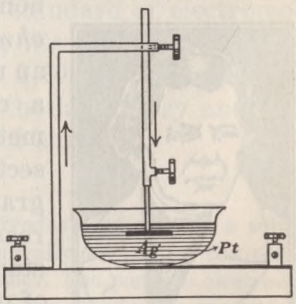


Fig. 328.—Silver coulometer

328. This is an electrolytic cell, consisting of a platinum bowl, which serves the double purpose of containing the electrolyte and acting as the cathode. The anode is a piece of pure silver. The electrolyte employed is a solution of silver nitrate, prepared according to standard specifications. The direction of the current is indicated by the arrows. Silver is deposited on the platinum cathode, the quantity thus deposited being determined by weighing the cathode (platinum bowl) both before and after the experiment. We define the unit of current strength, the ampere, in terms of the quantity of silver deposited per second.

An *ampere* is that steady current which will deposit 0.001118 gram of silver in one second. A *coulomb* is the quantity of electricity conveyed by a current of one ampere flowing for one second.

The *electrochemical equivalent* of an element is the number of grams deposited by a current of one ampere in one second, that is, the number of grams per coulomb. The electrochemical equivalent of silver is 0.001118.

372. **The Ohm.** Every conductor offers resistance to a current. The unit of resistance is the ohm, named in honor of Georg Simon Ohm. The *ohm* is the resistance offered to an unvarying current of electricity by a column of mercury 106.3 centimeters in length, of uniform cross section, and having a mass of 14.4521 grams, at a temperature of 0° C. The reasons for selecting mercury as the standard metal in measuring the unit of resistance are (a) that it has a high resistance as compared with other metals, and (b) it can be obtained in a very pure state by distillation.



Ohm

Ohm discovered that for a given conductor the current strength is proportional to the e.m.f. This means that for a given conductor the resistance is independent of the strength of the current flowing through it, provided the temperature remains constant.

373. The Volt. The unit of potential difference and e.m.f. is called the volt. A *volt* is the electromotive force that will cause a current of 1 ampere to flow through a

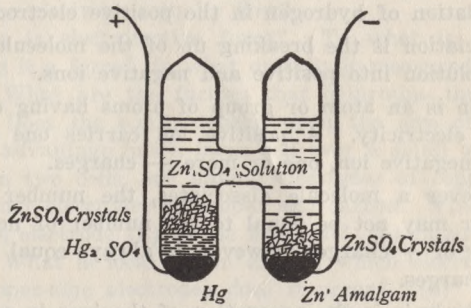


FIG. 329.—Weston standard cell

resistance of 1 ohm. The difference of potential applied to the ordinary incandescent lamp, for example, is 110 volts. The practical international standard of electromotive force is that furnished by the *Weston standard cell*, Fig. 329, the e.m.f. of which is 1.1083 volts at 20°C. This cell is easily reproducible in the laboratory and has a remarkably uniform and constant e.m.f.

SUMMARY

1. A voltaic cell consists of two electrodes dipping into a solution of such a nature that it acts chemically on one of the electrodes to a greater degree than on the other. The positive electrode is charged with electricity of the positive kind; the negative electrode, with electricity of the negative kind.

2. Electromotive force of a generator (battery or dynamo) is its capacity to generate electric pressure or difference of potential; e.m.f. is the cause of the flow of electric currents. The e.m.f. of a

battery is equal to the fall of potential around the entire circuit.

3. Electromotive force is not force; it is measured in units of work (ergs).

4. The e.m.f. of a cell depends upon the electrolyte, the kind of electrodes used, and the temperature; in a given cell it is independent of the size of plates or their distance apart.

5. Local action is the chemical action which takes place between an electrode and the impurities which it may contain.

6. Polarization is a lowering of the e.m.f. of a battery due to the accumulation of hydrogen in the positive electrode.

7. Dissociation is the breaking up of the molecules of a substance in solution into positive and negative ions.

8. An ion is an atom or group of atoms having one or more charges of electricity. A positive ion carries one or more + charges; a negative ion, one or more — charges.

9. Whenever a molecule dissociates, the number of positive ions may or may not be equal to the number of negative ions. The number of + charges, however, is always equal to the number of — charges.

10. Electrolysis is the separation of the ions in a solution by means of a current. The positive ions appear at the cathode; the negative ions at the anode.

11. Faraday's laws of electrolysis: I. The mass of a substance liberated in an electrolytic cell is proportional to the strength of the current and the time which the current flows. II. If the same quantity of electricity be passed through a number of electrolytic cells, each containing a different electrolyte, the mass of each substance deposited is proportional to its chemical equivalent.

12. An ampere is that current which will deposit 0.001118 gram of silver in a standard electrolytic cell in one second.

13. A coulomb is the quantity of electricity conveyed by a current of one ampere flowing for one second.

14. An ohm is the resistance of a standard column of mercury 106.3 cm in length and having a mass of 14.4521 grams, at 0° C.

15. A volt is that e.m.f. that will cause a current of one ampere to flow through a resistance of one ohm. The e.m.f. of a Weston standard cell is 1.1083 volts, at 20° C.

REVIEW EXERCISES

A

1. Make a sketch of a voltaic cell, and in connection with this sketch point out and define (a) positive electrode, (b) negative electrode, (c) electrolyte, (d) external circuit, (e) internal circuit.

2. Explain how we may determine which electrode of a cell has a positive charge and which a negative.

3. (a) In what direction in the external circuit is the current represented as flowing? (b) From what electrode to what electrode do the electrons actually flow?

4. What is electromotive force? To what is it sometimes likened? Is it a force? In what units is it measured?

5. (a) What are the factors that determine the e.m.f. of a cell? (b) Does the e.m.f. depend upon the size of the plates? (c) What advantage has a large cell over a small one?

6. Given two cells, one having electrodes of copper and zinc, and the other electrodes of copper and carbon. Which element is the positive electrode (a) in the first cell? (b) in the second cell?

7. (a) What is local action and at which electrode in a cell having copper-zinc electrodes does it occur? (b) Why is local action detrimental in a battery? (c) How may local action be remedied?

8. (a) What is polarization, and at which electrode (+ or —) does it occur? (b) Why is polarization in a cell detrimental? (c) How does it affect the current? (d) How may polarization be remedied?

9. Name the electrodes and the electrolytes used in each case of the following batteries: (a) gravity, (b) Leclanché. (c) Give the e.m.f. in each case. (d) Which of the two would you use on closed circuit work, and why?

10. (a) Describe the dry battery. (b) What is its e.m.f.? (c) How much electricity in ampere-hours should a fresh dry battery of standard size furnish?

11. (a) What is electrolytic dissociation? (b) What is an ion? (c) What is the difference between an ion and an electron?

12. Make a sketch of an electrolytic cell the electrolyte of which is a solution of HCl in water. Indicate (a) the direction of the current through the cell, (b) the direction in which the + ions move; (c) the — ions.

13. Explain briefly the decomposition of water by electrolysis.

What gas appears at (a) the anode? (b) the cathode? (c) What are the relative volumes of hydrogen and oxygen?

14. State Faraday's laws of electrolysis.

15. A current of one ampere will deposit by electrolysis 0.001118 gram of silver in one second. What current will be required to deposit 100 grams of silver in 2 hours?

16. In electroplating, what metal is used as the anode? Why?

17. (a) In charging a storage battery, do we put current into it or through it? (b) What is stored up in a charged battery?

18. (a) When a lead storage battery is charged, what is the color of the + electrode? (b) What is its chemical composition?

19. In charging a storage battery the positive terminal of the charging current should be connected with which pole of the battery, the one marked + or the one marked —? Why?

20. What is the e.m.f. of a lead storage battery when fully charged?

21. What is meant by the statement that the efficiency of a storage battery is 80 per cent?

22. State precautions to be observed in use of the lead storage battery.

23. What is an ampere, and how is its value determined experimentally? In the standard silver coulometer, what metal is used (a) as the anode? (b) the cathode? (c) What is the electrolyte?

24. What is a coulomb? Why is the electrolytic cell used in the experimental determination of the ampere called a silver coulometer?

25. What is an ohm, and how is its value determined experimentally? Why is mercury used?

26. Define the volt in terms of the ampere and ohm. What is the e.m.f. of a Weston standard cell?

B

27. Explain the distinction between electromotive force and difference of potential.

28. Write equations to illustrate the dissociation of (a) hydrochloric acid, (b) sulphuric acid, (c) copper sulphate. (d) How many positive and how many negative ions in each case?

29. (a) When a molecule dissociates, is the number of positive ions necessarily equal to the number of negative ions? (b) Is the

number of positive charges always equal to the number of negative charges?

30. By means of a sketch explain why the zinc electrode of a voltaic cell acquires a — charge and the copper electrode a + charge.

C

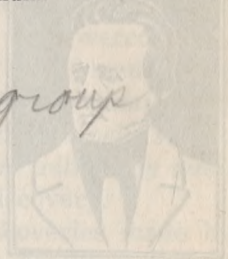
31. A current of one ampere flows through two electrolytic cells in series, one containing silver nitrate ($AgNO_3$) and the other copper sulphate ($CuSO_4$). (a) How much silver will be deposited in 1 hour? (b) How much copper?

32. (a) Why does the specific gravity of the electrolyte of a storage battery decrease on discharge and increase on charge? (b) What is the sp.g. value for full charge?

33. Describe briefly the Edison storage cell. Compare the nickel-iron cell with the lead cell.

NOTE: For additional problems, see Appendix.

write four in each group



CHAPTER XII

MAGNETIC AND HEATING EFFECTS OF A CURRENT

Magnetic Effects of a Current

374. The Discoveries of Oersted and Henry. It was early known that there was some very close relationship between electricity and magnetism. For example, it had been observed from time to time that a discharge of lightning sometimes had the effect of magnetizing knives and other steel objects. It was believed for a long time that this magnetizing effect was due to the passage of an electric current through the objects so affected. The true relationship between electrical and magnetic phenomena, however, was not discovered until 1819, when Hans Christian Oersted, a Danish physicist, who was conducting some experiments before his class in the University of Copenhagen, by chance placed a wire carrying a current near a magnetic needle; the needle was at once deflected from the north-south direction. When the conductor was parallel with the needle and above it, the N-pole was deflected in one direction; and when the wire was placed below the needle, the N-pole was deflected in the opposite direction. This effect of a current upon a magnetic needle is known as *Oersted's discovery.*



Oersted

Oersted's discovery aroused great interest among phy-

sicists, because it was at once perceived that it would not only be possible to signal at a distance by means of an electric current, but also that by means of a combination of current and magnetic systems it would be possible to transform the energy of an electric current into the energy of mechanical motion. One difficulty, however, stood in the way of the accomplishment of these highly desirable results, and that was the fact that for any considerable distance the resistance of the conductor was so great that the current flowing through it was reduced to such a low value as to make practical results impossible. This difficulty, however, was overcome by Joseph Henry, an American scientist, who invented the electromagnet. *Henry found that by winding many turns of insulated wire on an iron rod that relatively strong magnetic effects could be produced even with very weak currents.* This is known as Henry's discovery.



Joseph Henry

The results growing out of the discoveries made by Oersted and Henry proved to be far reaching and of great practical importance. Indeed, every piece of electromagnetic apparatus in use today, as for example, the electric bell, the telephone, the telegraph, and all forms of dynamo electric machinery include in one form or another an application of the principles involved in these epoch making discoveries. We shall now study in detail the magnetic effects of a current, and some of the practical applications of these effects.

375. Magnetic Field about a Current. That there exists a magnetic field about a wire carrying a current may be shown in a very striking manner as follows: Pass a wire through a hole in a piece of cardboard or a glass

plate upon which iron filings have been sprinkled. Now if the wire be connected to a source of e.m.f. and a rather strong current be passed through the circuit, the iron filings will arrange themselves in concentric lines about the condutor, thus showing that there is a magnetic field about the wire due to the current. The direction of the lines of force about the wire may be determined by placing a magnetic needle upon the plate, Fig. 330. If the current through the wire be reversed, the direction of the lines of force will be reversed, as shown by the reversal of the needle. *The relation between the direction of the current and the direction of the magnetic lines of force may be illustrated by grasping the wire with the right hand with the thumb in the direction of the current, in which case the fingers will represent the direction of the lines of force.*



FIG. 330.—Magnetic field about a wire carrying a current

376. Effects of a Current on the Magnetic Needle. A magnetic needle always tends to set itself parallel to the direction of the lines of force of the magnetic field in which it is placed, in such a way that the lines may be conceived of as entering the S-pole and coming out the N-pole. When a magnetic needle is placed in the earth's field, it therefore takes a north-south position. Now if a wire carrying a current be brought near the needle and above it, with a current flowing in the direction indicated in Fig. 331, the needle will be acted upon by two magnetic fields, the earth's field in one direction and the magnetic field due to the wire in a direction at right angles to the earth's field. Two forces therefore act upon the needle, which takes a position representing the direction of the resultant of these forces. When the wire is *above*

the needle, as shown in Fig. 331, the N-pole is deflected toward the observer. When the wire is placed *below* the needle, the current remaining in the same direction, the N-pole is deflected away from the observer.

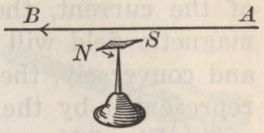


FIG. 331.—Deflection of a magnetic needle by a current

The right hand rule may conveniently be used to determine the direction and sense of the magnetic field around a conductor, and therefore the direction of deflection of the N-pole of a magnetic needle. Grasp the conductor with the right hand, the *thumb* pointing in the direction of the current, and the *fingers will indicate the direction of the magnetic lines of force* about the wire, Fig. 332. The deflection of the N-pole therefore will always be in the direction in which the fingers point.



FIG. 332.—Thumb indicates direction of current; fingers, direction of magnetic field

According to the right hand rule, the lines of force may be represented in another way, as illustrated in Figs. 333 and 334. Suppose we are looking at the conductor "end on"; that is to say, suppose that the small circle of Fig. 333 represents a circular conductor (wire carrying a current away from the observer), and Fig. 334 represents a conductor carrying a current toward the observer. The cross is the conventional symbol for an *in* current, and the dot represents an *out* current. The cross is said to symbolize the feathered end of an arrow going in, and the dot the point of the arrow coming out.

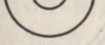


FIG. 333 "In" current

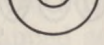


FIG. 334 "Out" current

Now consider Fig. 333. If we grasp the *in* conductor

with the right hand, the thumb pointing in the direction of the current, the direction of the lines and of the magnetic field will be in a clock-wise sense, as shown; and conversely, the magnetic field of the *out* current is represented by the curved line in a counter clock-wise sense, Fig. 334.

EXERCISES. **1.** Consider Fig. 331 with reference to the direction of current and the position of needle. Make drawings to illustrate the following cases, using the right hand rule to determine the deflection of the N-pole in each case: (a) Current in direction *BA* and above needle; (b) current in direction *BA* and below the needle.

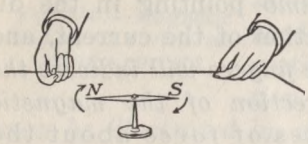


FIG. 335

2. Consider that the wire carrying the current, Fig. 331, is placed beside the needle, the current flowing from *B* to *A*. Determine by the right hand rule the effect on the needle.

3. Determine the direction of the current in the case of Fig. 335.

377. The Solenoid. A *solenoid* consists of a coil of wire of several turns, Fig. 336, carrying a current. When a

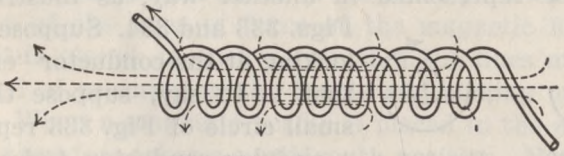


FIG. 336.—Solenoid

current passes through the solenoïd, the coil acquires the properties of a magnet, one face or end being an N-pole and the other face or end an S-pole.

The polarity of a solenoid may be shown as follows: Pass a current through several turns of insulated wire, Fig. 337. Present one end of the solenoid to the pole of a magnetic needle; the pole is deflected in a given direc-

tion. Now present the opposite end of the solenoid to the same pole; it is deflected in the opposite direction. Thus it appears that the solenoid, like the magnet, has two poles, N and S.

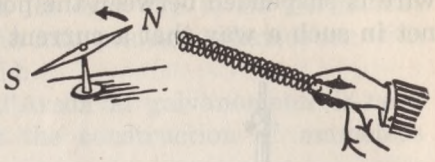


FIG. 337.—Experiment to determine the polarity of a solenoid

A very simple and convenient rule for determining the polarity of a solenoid may be stated thus: Grasp the solenoid with the right hand, the *fingers* pointing in the direction of the current, and the *thumb will indicate the direction of the N-pole,* Fig. 338.

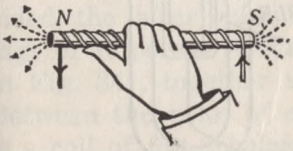


FIG. 338.—Right hand rule for determining the polarity of a solenoid

The solenoid principle is made use of in the galvanometer, the ammeter, and the voltmeter.

378. The Galvanometer. A *galvanometer* is an instrument for measuring small currents of electricity. One of the simplest forms is shown in Fig. 339. A small magnetic compass is placed in a wooden block or other non-magnetic support, around which is wrapped several turns of insulated wire. When a current flows through the conductor, the magnetic needle within the coil is deflected. This instrument is called a *tangent galvanometer* because the strength of current is proportional to the tangent of the angle of deflection. This form of galvanometer is not at present used very extensively.

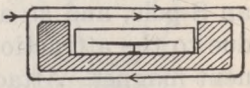


FIG. 339
Tangent galvanometer

A much more commonly used and convenient type of instrument is the *d'Arsonval galvanometer,* Fig. 340. The general principle upon which the instrument is based is

shown in the accompanying outline sketch. A coil of fine wire is suspended between the poles of a horseshoe magnet in such a way that a current may enter at one point

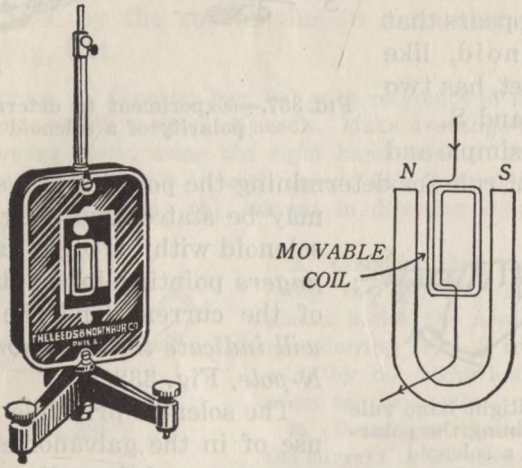


FIG. 340.—D'Arsonval galvanometer

and leave at another. Now when the current flows through the coil, one face becomes an N-pole, the other an S-pole, and as a result the coil swings about its axis due to the attraction of its poles for those of the permanent magnet. Attached to the coil is a pointer, or better still, a small mirror reflecting a beam of light. Thus, whenever a current flows through the instrument, its presence is indicated by the deflection of the coil. The great advantage of the d'Arsonval galvanometer over the tangent galvanometer lies not only in the fact that the former is more compact in construction and convenient in form than is the tangent galvanometer, but also that it is independent of the earth's field. The tangent galvanometer has to be set up in such a manner that the plane of the coil lies in the north-south direction. The

d'Arsonval instrument, on the other hand, having a very strong magnetic field of its own due to its permanent magnet, may be set up in any position whatsoever with respect to the earth's field.

The principle of the d'Arsonval galvanometer is very extensively employed in the construction of ammeters and voltmeters.

379. The Ammeter. An *ammeter* is an instrument for measuring current strength in amperes. It is really a galvanometer, usually of the d'Arsonval type, having a coil of very low resistance. A pointer and scale is provided, the latter being calibrated so as to indicate readings in amperes. A standard type of ammeter is shown in Fig. 341, together with its principal working parts. Between the poles of a strong horseshoe magnet there is a coil of fine insulated wire wound on a light frame,

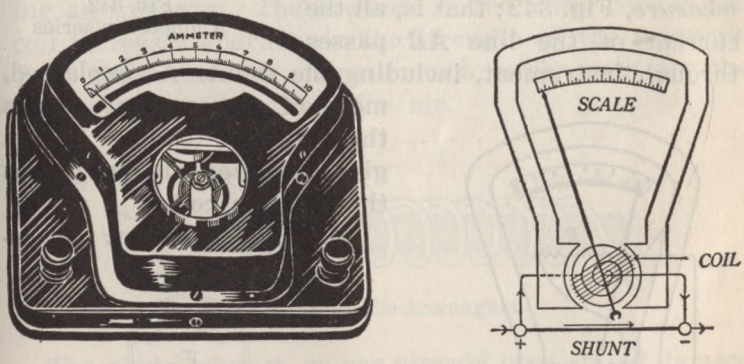


FIG. 341.—Ammeter

the axis of which is supported upon jewelled bearings. Attached to the movable coil there is a long pointer which moves over a scale calibrated to give readings in amperes. By means of two spiral springs, one at each end of the frame upon which the coil is wound, the coil is held in equilibrium so that when no current is passing

through the instrument, the pointer rests at the zero reading of the scale.

Since the moving coil of an ammeter is too delicate to carry large currents, most of the current which enters the instrument is conveyed from the positive terminal to the negative by means of a *shunt*, which is a short heavy wire of low resistance. Usually about 0.01 of the current which enters the ammeter passes through the movable coil, while 0.99 pass by way of the shunt. The scale is so calibrated that it indicates the total current which is passing from terminal to terminal and not the small fraction which passes through the movable coil.

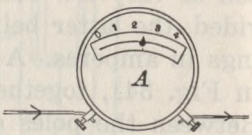


FIG. 342
Ammeter in series

An ammeter is always put in series with the circuit, the current of which it is desired to measure, Fig. 342; that is, all the current of the line *AB* passes through instrument, including the shunt. As explained, most of the current passes through the shunt, but a given fraction always passes through the coil.

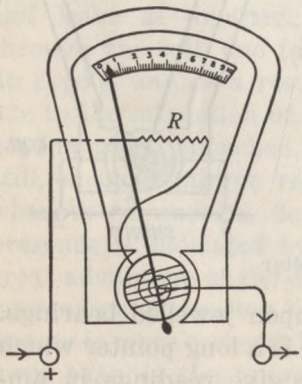


FIG. 343.—Voltmeter

380. The Voltmeter. A *voltmeter* is an instrument for measuring the difference of potential in volts, Fig. 343. Like the ammeter it is also

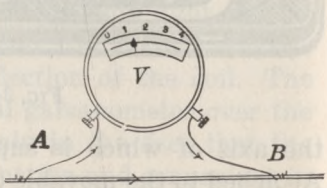


FIG. 344
Voltmeter in shunt

a d'Arsonval galvanometer, but differs from the ammeter in three respects: first, it has no shunt; second, it has a high resistance R in series with the movable coil; and third, it is calibrated to give readings in volts.

A voltmeter is always put in parallel or shunt with the conductor over which the potential difference is to be measured, Fig. 344, and it is for this reason that voltmeters are constructed with coils of high resistance; that is, so that the instrument will draw only a relatively small current from the main circuit. The voltmeter shown in Fig. 344 measures the fall of potential in volts between the points A and B.

381. The Electromagnet. An *electromagnet* consists of a soft iron bar about which there is wound a number of turns of insulated wire, Fig. 345. The electromagnet, it will be observed, is nothing more than a solenoid having an iron core. The presence of the metal core in the coil increases enormously the strength of the magnetic field, since the iron offers a much easier path for the lines of force than does the air.

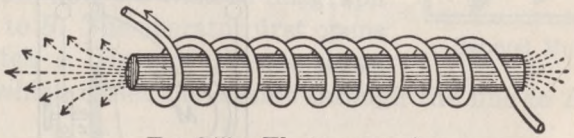


FIG. 345.—Electromagnet

The electromagnet, as has already been stated, is used as an essential part of such electromagnetic pieces of apparatus as the electric bell, the telephone, the telegraph, the dynamo, etc. Powerful electromagnets are employed in the loading and unloading of iron, as shown in Fig. 346. When the current is turned on, the electromagnet is capable of lifting a ton or more of iron. The crane to which the magnet is attached is then swung

into any given position where it is desired to drop the material and the current is cut off, allowing the load B to fall. The object of using soft iron as the core of an electromagnet is that soft iron becomes strongly magnetized when the current is on, and quickly loses its magnetism when the current is cut off. Tempered steel would therefore not do to use as the core of such a magnet because of its tendency to become permanently magnetized. The current is supplied to the magnet M through the line L.

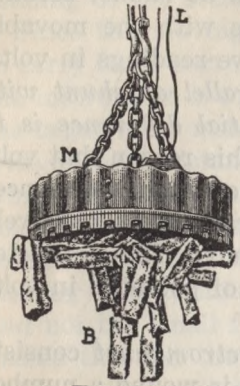


FIG. 346
Powerful electromagnet used for unloading iron bars

Applications of the electromagnet are found in the electric bell and the telegraph instrument.

382. The Electric Bell. The essential parts of an electric bell system are the battery, the push button, and

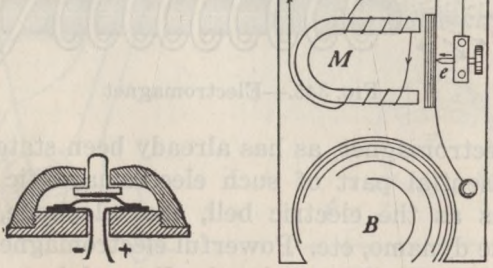


FIG. 347.—Push button and electric bell

the bell. The push button is a device for making and breaking the circuit. The parts of an electric bell are

shown in outline in Fig. 347. When the circuit is closed by pressing the button, the current from the battery magnetizes the soft iron core of the electromagnet *M*. This attracts the iron strip or armature, which is held in place by a light spring. When the armature is pulled down by the electromagnet *M*, the circuit is broken at *e*. As the armature moves downward, it causes the clapper to strike the bell *B*. The moment the circuit is broken, however, the spring throws the armature back to the point *e*, thus again closing the circuit.

383. The Telegraph. The electric telegraph was invented by Samuel Morse in 1832. The operation of telegraphing from one point to another by means of the *key* and *sounder* system may be explained in connection with Fig. 348. The current from the battery at station *A* may be traced through the electromagnet *m*, the line *L*, the electromagnet *m'*, the battery at *B*, thence through the earth back to *A*.



Samuel Morse

Suppose now we wish to telegraph from *A* to *B*. The operator first opens his switch *s*. He now presses down key *k*, which closes the circuit through the line to *B* and

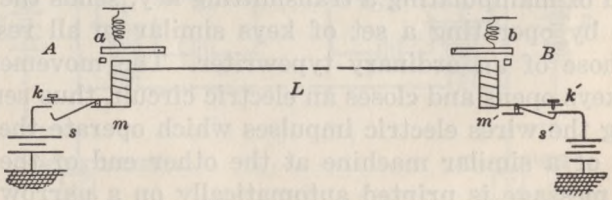


FIG. 348.—Diagram of simple telegraph

back through the earth. When the current passes through the two electromagnets *m* and *m'*, *a* and *b* are drawn

down, giving rise to the clicking noise characteristic of the telegraph instrument. The combination of magnet and armature is called a sounder. Every time the key at *A* is pressed down there is produced in the sounder at both *A* and *B* corresponding long or short clicking intervals called dots and dashes. The operator at *B* reads the dots and dashes as they are sounded upon his instrument, and writes out the message. When the operator at *A* is through sending, he closes the switch *s*. This diagram above shows only two stations. In actual practice a great many stations may be on the line. Key and sounder telegraph instruments are shown in Fig. 349.

The old fashioned receiving telegraph instruments have been replaced within the last few years on all the

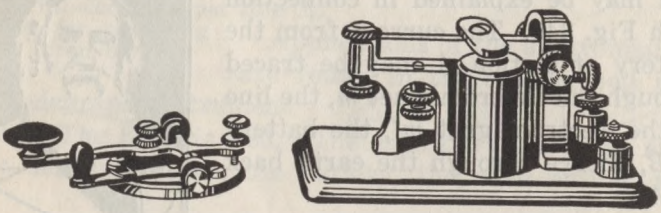


FIG. 349.—Telegraph key and sounder

large telegraph lines by the *tape-teletype*, Fig. 350, which is essentially an electric typewriter. The operator, instead of manipulating a transmitting key, sends the message by operating a set of keys similar in all respects to those of an ordinary typewriter. The movement of the keys opens and closes an electric circuit, thus sending along the wires electric impulses which operate the type bars of a similar machine at the other end of the line. The message is printed automatically on a narrow tape at the receiving end. All the receiving operator has to do is to paste the gummed type on the telegram sheet as the printed slip comes from the machine.

The picture shown in Fig. 350 is that of a sending-receiving model. In some large stations the sending and

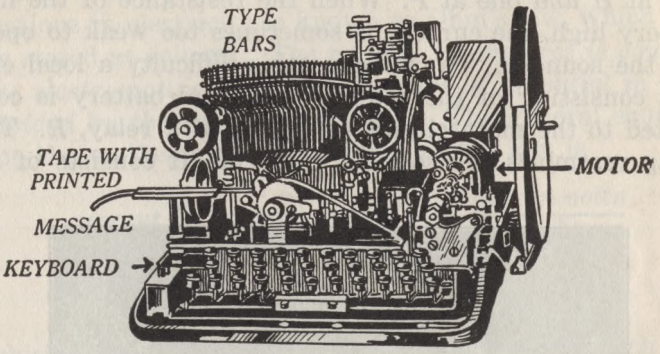


FIG. 350.—Tape-teletype

receiving units are mounted as separate machines. Each tape-teletype set is equipped with a small electric motor for driving the various mechanisms. Standard sets can be arranged for operation on either direct or alternating currents.

384. Telegraph and Cable Lines. In commercial long distance telegraph systems currents from storage batteries or dynamos are used in the main line while in the

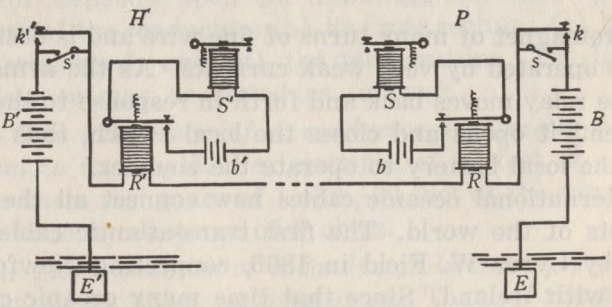


FIG. 351.—Telegraph line showing sounders and relays

local circuit voltaic cells, usually of the gravity battery type, are employed.

A diagram of a complete line and local telegraph system is shown in Fig. 351. Two stations are represented, one at *H* and one at *P*. When the resistance of the line is very high, the current is sometimes too weak to operate the sounder. To overcome this difficulty a local circuit consisting of the sounder and local battery is connected to the main circuit by means of a relay, *R*. The *relay* resembles the sounder in form. It consists of an

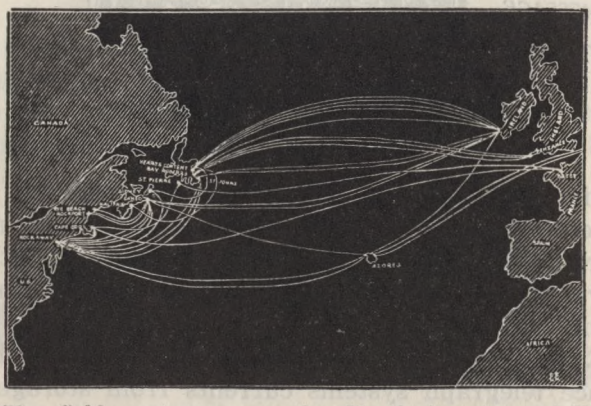


FIG. 352.—Cable systems connecting North America and Europe

electromagnet of many turns of fine wire and is designed to be operated by very weak currents. As the armature of the relay moves back and forth in response to the line current, it opens and closes the local switch, thus causing the local battery to operate the sounder.

International oceanic cables now connect all the continents of the world. The first trans-atlantic cable was laid by Cyrus W. Field in 1866, connecting Newfoundland with Ireland. Since that time many oceanic cables have been installed and are now in successful operation. A diagram of the Atlantic cables connecting North America with Europe is given in Fig. 352.

OHM'S LAW AND ITS APPLICATION

385. Ohm's Law. One of the most important generalizations in electricity is known as Ohm's law, which may be stated as follows: *The current in amperes is equal to the electromotive force or difference of potential in volts divided by the resistance in ohms.* The general equation for the mathematical expression of the law is

$$\textit{current in amperes} = \frac{\textit{e.m.f. or p.d. in volts}}{\textit{resistance in ohms}}$$

$$I = \frac{E}{R}$$

which implies that the current in any system, such as a battery and its conducting circuit, is *directly* proportional to the electromotive force and *inversely* proportional to the resistance. This means, in the case of a battery system having a given resistance, the greater the e.m.f. the greater the current; and on the other hand, for a given e.m.f. the greater the resistance the less the current.

386. The Laws of Resistance. The resistance of a conductor depends upon the following factors: (a) the length of the conductor; (b) its cross section; (c) kind of material used; and (d) the temperature. The laws of resistance may be stated as follows:

I. *The resistance of a conductor is directly proportional to its length.* For example, if 10 feet of a given wire has a resistance of 1 ohm, 20 feet of the same wire will have a resistance of 2 ohms.

II. *The resistance of a conductor is inversely proportional to its cross sectional area.* That is, the greater the cross sectional area of a conductor, the less the resistance. Since the cross sectional area of a wire is proportional to the square of its diameter d, it is sometimes con-

venient to consider the resistance as inversely proportional to the square of the diameter, d^2.

III. *The resistance of a conductor depends on the material of which it is composed.* The resistance of iron, for example, is more than six times as great as that of copper, for the same length and cross section.

IV. *The resistance of a metallic conductor increases as the temperature increases.* The hotter a wire becomes, the greater is its resistance. On the other hand, *the resistance of carbon and electrolytes decreases as the temperature increases.* The resistance of the carbon filament of a 110 volt 16 candle power incandescent lamp is, when cold, about 440 ohms; when hot (incandescent) about 220 ohms.

387. Discussion of the Laws of Resistance. The first three laws of resistance may be expressed in the form of equations as follows:

$$R = \frac{k \times l}{a}$$

$$= \frac{k' \times l'}{d^2}$$

In the first equation given above R is the resistance in ohms, l the length of the conductor in *centimeters*, a the cross sectional area of the conductor in *square centimeters*, and k is a constant depending upon the nature of the material of which the conductor is composed. The constant k in the equation is expressed in ohm-centimeters, as given in Table XVII, Appendix.

In the second equation R is again the resistance in ohms, l' the length in *feet*, d is the diameter of the wire, usually expressed in thousandths of on inch, and k' is a constant, which is expressed in *ohms per mil-foot*. A *mil* is a thousandth (0.001) of an inch; or to put it another way, 1 inch = 1000 mils. Values of k' in ohms per

mil-foot are given in Table XVII, Appendix. The square of the diameter of a wire in thousandths of an inch is sometimes called by engineers a *circular mil* (*C.M.*), that is $d^2 = C.M.$

Example. 1. Find the resistance at 0° C of 5 *m* of copper wire, the radius of which is 0.5 *mm*. *Solution*: On referring to Resistivity Table XVII, Appendix, we find that the value of k for copper in ohms per *cm* is $1.6 \times 10^{-6} = 0.0000016$. The length $l = 5\ m = 500\ cm$. Also, 0.5 *mm* = 0.05 *cm*, and therefore $a = \pi r^2 = \pi \times (0.05)^2 = \pi \times 0.0025$. Then

$$R = \frac{k \times l}{a}$$

$$= \frac{0.0000016 \times 500}{\pi \times 0.0025} \qquad \frac{0.32}{\pi} \text{ ohm}$$

Example 2. Find the resistance at 0° C of 1 mile of copper wire the diameter of which is 0.03 inch. *Solution*: The value of k' for copper, Table XVII, Appendix, in ohms per mil-foot is 10.1. The length l' = 1 mile = 5280 ft. The diameter d = 0.03 inch. Since 1 inch = 1000 mils, 0.03 inch equals $0.03 \times 1000 = 30$ mils, and hence $d^2 = 900$. Then

$$R = \frac{k' \times l'}{d^2}$$

$$= \frac{10.1 \times 5280}{900} = 59.25 \text{ ohms}$$

Example 3. Find the resistance of the copper conductors mentioned in examples 1 and 2 at room temperature (20° C). *Solution*: In general, the resistance of metallic conductors increases from 0° C values at the rate of 0.004 ohm per degree. The respective resistances then for 20° C are:

(a) Ex. 1, $R = \frac{0.32}{\pi} + \left(\frac{0.32}{\pi} \times 0.004 \times 20\right) = \frac{0.346}{\pi}$

(b) Ex. 2, $R = 59.25 + (59.25 \times 0.004 \times 20) = 63.99$ ohms.

388. The Resistance Box. In measuring resistance it is necessary to have at hand a known resistance, such as is furnished by the *resistance box* shown in Fig. 353.

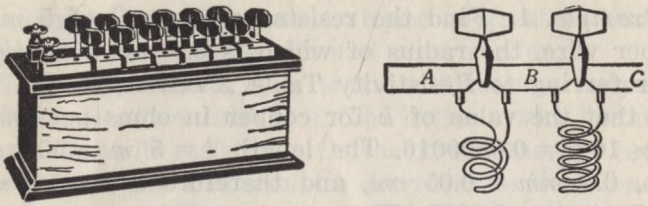


FIG. 353.—Resistance box and resistance coils

Such a box contains a set of resistance coils. A resistance coil is made by winding a double wire in the form of a spiral, the two free ends being fastened to the metal plates *A* and *B*, *B* and *C*, which in turn are set in the resistance box. The object of winding the wire double is to prevent magnetizing effects when the current passes through the coil, and also to prevent self-induction, as will be explained in Art. 415. The best modern resistance coils are made of *magnanin* wire, the resistance of which is practically unaffected by ordinary changes of temperature such as occur in the laboratory. These coils are mounted in boxes with their terminals connected to the brass plates as shown. Connection from one wire of the coil to the next is made by means of a metal plug. When the plugs are all in place, the resistance of the box is practically zero. When a plug is withdrawn from the box, the resistance offered is equal to that of the coil beneath the particular plug withdrawn.

Resistance boxes may be bought with coils varying in resistance from 0.1 to 1000 ohms, and even greater ranges.

389. Resistance of Conductors in Series. When conductors are joined end to end as shown in Fig. 354, they

are said to be connected in *series*. The resistance of a

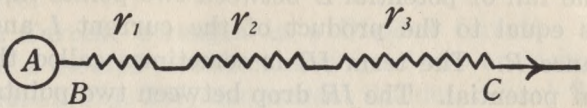


FIG. 354.—Conductors in series

number of conductors connected in series is equal to the sum of the individual resistances; that is

$$R = r_1 + r_2 + r_3 + \ldots$$

Example. Three wires having resistances of 5, 10, and 15 ohms respectively are connected in series. Find the total resistance of the combination. *Solution*: $R = 5 + 10 + 15 = 30$ ohms.

390. Resistance of Conductors in Parallel. When two or more conductors are connected as shown in Fig. 355,

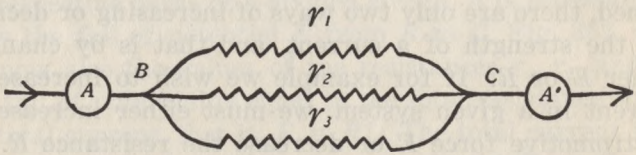


FIG. 355.—Conductors in parallel

they are said to be in *parallel*. It may be demonstrated that the resistance R of two or more conductors in parallel may be expressed by an equation of the following form:

$$\frac{1}{R} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3} \cdots$$

Example. Let r_1 Fig. 355, be a resistance of 5 ohms, r_2 10 ohms, and r_3 15 ohms. Find the resistance of the three wires thus connected in parallel. *Solution*: $1/R = 1/5 + 1/10 + 1/15$, hence $R = 2.7$ ohms.

391. Discussion of Ohm's Law. The equation for

Ohm's law, $I = E/R$, may be written $E = IR$. This means that the fall of potential E between two points in a circuit is equal to the product of the current I and the resistance R. The term IR is sometimes called the *IR drop* of potential. The IR drop between two points in a circuit carrying a current is the difference of potential between these points; that is, IR drop = p.d in volts.

Example. The resistances shown in Fig. 354 are as follows: $r_1 = 5$ ohms, $r_2 = 10$ ohms and $r_3 = 15$ ohms. The current flowing through the system, as indicated by the ammeter A, is 2 amperes. Find (a) the IR drop over each resistance, and (b) the total p.d. between B and C. *Solution*: (a) $Ir_1 = 2 \times 5 = 10$ volts, $Ir_2 = 20$ volts, and $Ir_3 = 30$ volts. (b) The total p.d. between B and C is the sum of $Ir_1 + Ir_2 + Ir_3 = 2 \times (5 + 10 + 15) = 60$ volts.

392. Current Strength in a Series Circuit. We learn from Ohm's law that, so far as direct currents are concerned, there are only two ways of increasing or decreasing the strength of a current, and that is by changing either E or R. If for example we wish to increase the current in a given system, we must either increase the electromotive force E or decrease the resistance R.

It is important to note that when the value of E and R are once fixed in a *series circuit*, the current strength is then the same throughout all parts of the circuit. Thus, if in a circuit containing the resistances r_1 and r_2, Fig. 356, an ammeter be introduced successively at A, A', and A'', it will be found that the reading of the instrument (current in amperes) is the same at all points; that is, if the current strength at A is 3

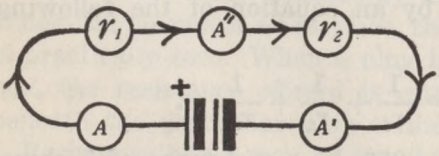


FIG. 356.—Current of same strength throughout the circuit

amperes, the current strength at A' and A'' is also 3 amperes.

393. Current Strength in a Divided Circuit. In the case of a *divided circuit* consisting of two or more parts in parallel, Fig. 357, the current divides, a part flowing over each path. The current divides at B and the fractional currents unite again at C, as is demonstrated by the fact that the reading of the ammeter at C is equal to that at B.

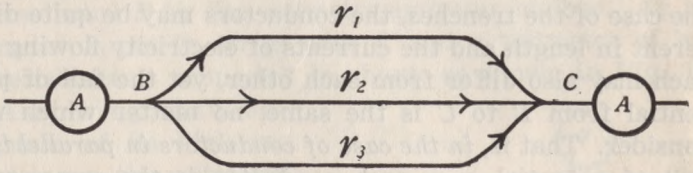


FIG. 357.—Divided circuit

In order to find the values of the current flowing over the respective paths of a divided circuit, it is necessary to know the fall of potential E from B to C, Fig. 357, and to know also the values of the resistances r_1, r_2, and r_3. The fall of potential is equal to the total resistance times the total current, that is, $E = RI$. The total current may be determined by reading the ammeter at B or C; the total resistance of the branches may be determined from the equation

$$\frac{1}{R} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3}$$

Now having determined E over the divided circuit we may compute the current in the various branches thus:

$$I_1 = \frac{E}{r_1},\ I_2 = \frac{E}{r_2},\ I_3 = \frac{E}{r_3}$$

It should be noted that the total current as recorded by the ammeter is $I = I_1 + I_2 + I_3$.

394. Fall of Potential over Conductors in Parallel. Suppose that water flows from a point B through three

trenches to a point C on a lower level. Now while the trenches extending from B to C may be of different lengths and the water flowing in any trench may be entirely different in quantity from that of any of the other trenches, yet we readily comprehend that the difference of pressure between the two points B and C is the same, no matter what path we consider. Suppose now that a current from a point B at high potential flows over three paths or conductors to a point C of low potential. As in the case of the trenches, the conductors may be quite different in length and the currents of electricity flowing in each may also differ from each other, yet the fall of potential from B to C is the same, no matter which we consider. That is, *in the case of conductors in parallel the fall of potential over each conductor is the same,* and furthermore, the fall of potential over each conductor is the same as the fall of potential from B to C obtained by multiplying the total current by the total resistance between these points.

Example. Three conductors, 4, 6 and 12 ohms respectively, connect the points B and C, Fig. 357. A current of 6 amperes flows from B to C, a part passing through each conductor. Find (a) the fall of potential from B to C; (b) the current through each wire. *Solution*: (a) The current from B to $C = 6$ amperes; the resistance from B to $C = 2$ ohms. Therefore, the fall of potential from B to $C = 6 \times 2 = 12$ volts. (b) Now since the fall of potential from B to C, as just found, is 12 volts, the fall of potential over any of the three paths is also 12 volts. Hence the current in the first path will be $12/4 = 3$ amperes; the current in the second path $12/6 = 2$ amperes; the current in the third path $12/12 = 1$ ampere. The total current in all three paths $= 3 + 2 + 1 = 6$ amperes.

395. Resistance by Ohm's Law. The equation $E = IR$

suggests also a method of measuring the resistance of a conductor by the application of Ohm's law. We may write $R = E/I$; in order then to determine the value of R, we must know the values of E and I. To determine these values experimentally, it is necessary to use a voltmeter and an ammeter; for this reason the method is sometimes spoken of as the *voltmeter-ammeter method* of measuring resistance.

Suppose that we desire to find the resistance of the conductor R by the voltmeter-ammeter method. An ammeter is put in series with R and a voltmeter of high resistance is connected in shunt, as shown in Fig. 358. Assuming that, because of its high resistance, the voltmeter draws only a negligible current from the circuit, the current I flowing through R then is practically that registered by the ammeter. The voltmeter measures the fall of potential E across the terminals of the conductor R. Let us assume that the current strength as measured by the ammeter is 2 amperes, and that the fall of potential across R is 10 volts. The resistance of the conductor then is $R = 10/2 = 5$ ohms.

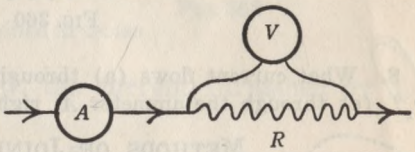


FIG. 358.—Resistance by the voltmeter-ammeter method

EXERCISES. **4.** Given two conductors, the resistance of r_1 being

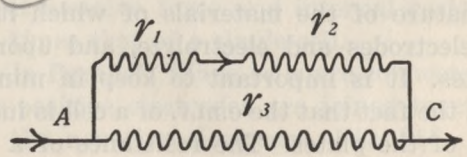


FIG. 359

10 ohms and that of r_2 15 ohms. Find the resistance R of these conductors when connected in parallel.

5. Three wires having a resistance of 2, 4, and 6 ohms respectively are connected in parallel. Find the resistance of the system.

6. The three wires of problem 5 are connected as shown in Fig. 359. Find the resistance of the system from A to C, when $r_1 = 2$, $r_2 = 4$, and $r_3 = 6$.

7. Two conductors, r_1 having a resistance of 2 ohms and r_2 a resistance of 8 ohms, are put in parallel, Fig. 360. Ammeter A, left, registers a current of 10 amperes. Find (a) the fall of potential over r_1, (b) the fall of potential over r_2.

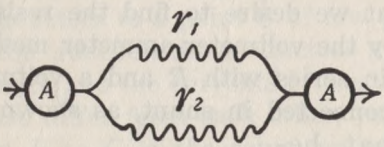


FIG. 360

8. What current flows (a) through r_1 (Ex. 7)? (b) through r_2? (c) through the ammeter A, right?

METHODS OF JOINING CELLS

396. Introductory. Ohm's law may be applied to determine the best method of joining the cells of a battery. The current furnished by a battery of one or more cells is determined by the electromotive force E and by the total resistance of the system. The resistance of a battery circuit is made up of two parts, that of the conductors connected to the poles, called the external resistance R, and that of the battery itself, called the internal resistance r. The internal resistance of a battery depends upon the nature of the materials of which it is made, including electrodes and electrolyte, and upon the size of the plates. It is important to keep in mind, in this connection, the fact that the e.m.f. of a cell is independent of the size of the plates. The resistance of a cell, however, diminishes as the size of the electrodes (plates) is increased; in other words, the larger the electrodes the less the internal resistance r.

The cells of a battery may in general be joined in two ways: (a) in series, and (b) in parallel.

397. Cells in Series. When cells are connected in *series,* the connecting wires join positive pole to negative pole, as shown in Fig. 361. A diagrammatic representation of cells in series is given in Fig. 362.

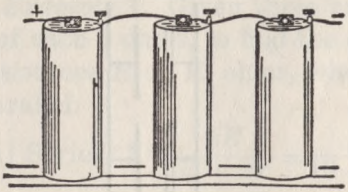


FIG. 361

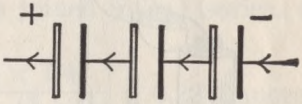


FIG. 362

Cells joined in series

According to Ohm's law, the current furnished by a *single cell* is

$$I = \frac{E}{R + r}$$

in which R is the resistance of the external circuit and r is the internal resistance of the cell.

When a number of cells (n) is joined in series, the equation becomes for *n cells in series*

$$I = \frac{nE}{R + nr}$$

Thus the effect of joining n cells in series is to make the total electromotive force and internal resistance respectively n times that of a single cell.

398. Cells in Parallel. When cells are connected in *parallel,* all the positive electrodes are joined together, and likewise all the negative electrodes, Fig. 363. The diagrammatic representation of cells in parallel is shown in Fig. 364. The effect of connecting a group of similar cells constituting a battery in parallel is equivalent to

producing a single cell, having an e.m.f. equal to one of the original cells, the plates of which are n times as large as those of any one of the original cells, and the internal resistance, therefore, is r/n. Since the electromotive force of a battery is independent of the size of the plates, the

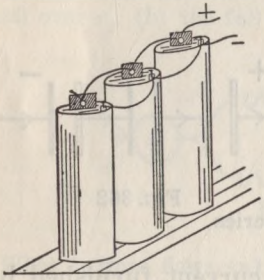


FIG. 363

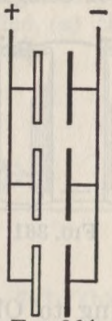


FIG. 364

Cells joined in parallel

e.m.f. of n cells joined in parallel is equal to that of any single cell. By Ohm's law, therefore, we may write for *n cells in parallel*

$$I = \frac{E}{R + r/n}$$

399. How to Join Cells for Maximum Current. We have just considered the joining of the cells of a battery in series and in parallel. The question naturally arises as to which is the better plan. This depends on the relation of the resistance of the external circuit as compared with the internal resistance of the battery. When the external resistance R is large as compared with the internal resistance r, the series plan gives the greater current. When R is small as compared with r, the parallel plan is the better.

In order to demonstrate the truth of these statements, let us take three cells of given e.m.f. and internal resist-

ance r and join them in series and in parallel, first with the external resistance R greater than r, and second with R less than r. Take for example three dry cells, the e.m.f. of each being 1.4 volts and the internal resistance of each 3 ohms. We shall join them in series and in parallel, first with R 10 ohms and second with R 0.1 ohm.

Example 1. Given three cells, e.m.f. of each 1.4 volts, r of each 3 ohms, to find the current through an external resistance R of 10 ohms, when joined in (a) series, (b) parallel.

(a) Series: $I = \dfrac{nE}{R + nr} = \dfrac{3 \times 1.4}{10 + (3 \times 3)} = 0.22$ ampere

(b) Parallel: $I = \dfrac{E}{R + r/n} = \dfrac{1.4}{10 + 3/3} = 0.13$ ampere

Example 2. Find the current furnished by the three cells, when the external R is 0.1 ohms, (a) series, (b) parallel.

(a) Series: $I = \dfrac{nE}{R + nr} = \dfrac{3 \times 1.4}{0.1 + (3 \times 3)} = 0.46$ ampere

(b) Parallel: $I = \dfrac{E}{R + r/n} = \dfrac{1.4}{0.1 + 3/3} = 1.27$ amperes

Conclusions: (a) When the external resistance R is greater than the internal resistance of a single cell, the series plan of connecting the cells of a battery gives the greater current. (b) When R is less than r, the parallel method is the more desirable. (c) Since the external resistance R is nearly always greater than the internal resistance r, cells are in general connected in series. The parallel method is sometimes employed, however, in charging and discharging storage batteries.

Electric Heat and Power

400. The Electric Current as a Heating Agent. Whenever a current of electricity flows through a conductor, heat is developed. The case that is most familiar is that

of the heating of the filament of an incandescent lamp. Other illustrations of the heating effect of a current are exemplified in the use of the electric flatiron, the electric soldering iron, and similar appliances. Electric currents are also used in some places for cooking, and occasionally for the heating of rooms. Electric ovens are now made which are automatic with reference both to time and temperature. The time control consists of an electrically driven time clock that turns the oven heat on at any given time and off at a later time. The "on" hand is set at the time to start the cooking, and the "off" hand at the time the cooking is to stop. Such insulated ovens will keep the food warm for an hour or more after the heat has been automatically turned off. Temperature con-

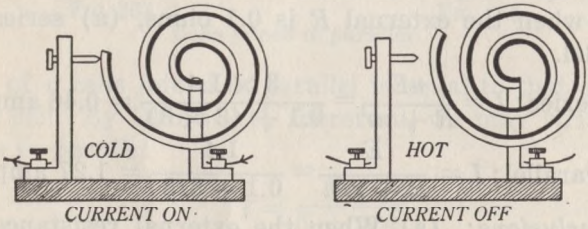


FIG. 365.—Oven thermostat

trol is obtained by means of an oven thermostat, the expanding element of which is a metal bar, Fig. 365. The thermostat may be set to keep the oven at any desired temperature for the time that it is needed.

The diagram shown in Fig. 366 illustrates the arrangement of the heating elements of the electric flat iron. These heating elements (conductors) are usually built up by winding a thin ribbon of "nichrome" (nickel-chromium) upon an insulating support made of asbestos or mica. In the use of the electric current as a heating agent in the cook stove or range, great care should be

exercised to keep the heating elements free from moisture or other conducting media to prevent the short circuiting of the heating elements.

Electric heating recommends itself more on account of its convenience than on account of its cheapness. In the

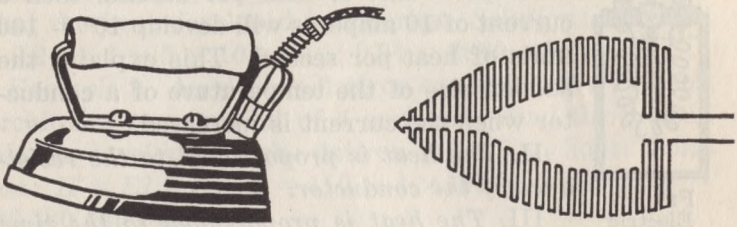


FIG. 366.—Electric flatiron and heating elements

case of an electric flatiron, for example, the heat may be turned on or off by simply opening or closing a switch. In this connection it should be remembered that the utmost care should be exercised to see that the circuit is opened when the instrument is not in use, as serious fires have been caused by over heated flatirons and other like pieces of domestic electrical heating appliances. When one is through with a flatiron, the best way to break the circuit is to detach the connecting plug from the iron; one is then sure that the current is broken and that no damage will result.

The power used by the ordinary electric flatiron is equivalent to that consumed by about 20 25-watt incandescent lamps.

401. Joule's Law of Electric Heating. By means of a simple calorimeter, Fig. 367, in which a current of electricity was passed through a wire immersed in a nonconducting fluid, Joule discovered that the heat developed is proportional to the square of the current, the resistance of the conductor, and the time which the current

flows. Joule's law of electric heating may for convenience be expressed in three parts as follows:

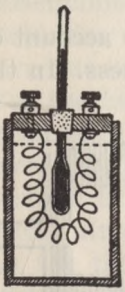


FIG. 367
Electric calorimeter

1. *The heat is proportional to the square of the current.* That is, if a current of 1 ampere flowing through a given conductor develops 1 unit of heat per second, then a current of 10 amperes will develop $10^2 = 100$ units of heat per second. This explains the sudden rise of the temperature of a conductor when the current is increased.

II. *The heat is proportional to the resistance of the conductor.*

III. *The heat is proportional to the time which the current flows.* Joule's law may be put in equational form, thus:

$$calories = amperes\ squared \times ohms \times seconds \times 0.24$$
$$H = I^2Rt \times 0.24$$

The factor 0.24 which appears in this equation was determined experimentally; it is the number by which it is necessary to multiply in order to express H in calories.

It is sometimes desirable to write the foregoing equation in a little different way so as to express the heat value in terms of volts times amperes. Since $I^2R = IR \times I$, and also since $IR = E$, we may write our general equation as follows:

$$H = EIt \times 0.24 = calories$$

where H, as before, is expressed in calories, E is the difference in potential over the conductor, measured in volts; I is the current in amperes, and t the time in seconds.

One B.t.u. is the heat required to change the temperature of 1 pound of water 1° F. But 1 pound = 453.6

grams and 1 F degree = 5/9 C degree, therefore 1 B.t.u. = $453.6 \times 5/9 = 252$ calories. Hence

$$B.t.u. = \frac{calories}{252}$$

Example 1. A current of 2 amperes flows through a conductor having a resistance of 5 ohms for 10 minutes. Find the heat developed in calories. *Solution*: $H = I^2Rt \times 0.24 = 4 \times 5 \times 10 \times 60 \times 0.24 = 2880$ calories.

Example 2. An electric flatiron connected to a 110 volt circuit draws a current of 4 amperes. Find the heat in calories developed in the flatiron in half an hour. *Solution*: $H = EI\,t \times 0.24 = 110 \times 4 \times 30 \times 60 \times 0.24 =$ 190,080 calories.

EXERCISES. **9.** A current of 3 amperes flows through a given wire conductor for 2 hours. The length of the wire is 4 m; its cross sectional area, 0.03 cm^2; the value of k for the given temperature, 0.0000174. Find the heat in calories developed in the conductor.

10. An electric toaster which takes 5 amperes at 110 volts toasts a slice of bread in 2 minutes. Find the heat developed in calories.

402. Electric Fuses. In the principle involved in the use of the safety fuse, advantage is taken of the heating effect of a current. The *electric fuse* is a device for automatically breaking the circuit in case, for a given system, the current becomes dangerously high. Fuse wires are made of some high resistance material which has a low melting point. Suppose, for example, that a fuse be placed in a circuit containing a piece of apparatus, the maximum carrying capacity of which is 2.5 amperes, but which is designed to be operated, under ordinary conditions, with a current of 2 amperes or less. If now for any reason the current should tend to rise above the carrying capacity of the instrument the fuse wire would melt at 2 amperes and thus save the apparatus from burning out.

Since the melting or "blowing" of a fuse wire is always accompanied by the formation of an electric arc which is likely to set fire to nearby inflammable material, fuses



FIG. 368

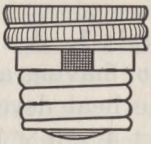


FIG. 369

Electric fuse plugs

are usually enclosed in fireproof receptacles called fuse plugs, Fig. 368.

In dwelling houses a capsule like fuse plug is frequently employed, Fig. 369, which is so designed as to screw into place in the fuse box. The capacity of these house fuses is usually from 10 to 15 amperes; that is, they are designed to melt when the current exceeds this value.

403. Electric Power. The power expended by an electric current (electrical energy per second) is expressed in watts, or more frequently in kilowatts. The *watt* as defined electrically is volts times amperes. Since $E = IR$ and $EI = I^2R$, therefore we may write

$$\begin{aligned} watts &= volts \times amperes \\ &= EI = I^2R \end{aligned}$$

In this connection it should be noted that a watt as defined electrically does not differ in magnitude from the watt as defined mechanically (Art. 181). That is to say, 1 volt $\times$ 1 ampere = 1 watt = 10,000,000 ergs per second = 1 joule per second. One *kilowatt* = 1000 *watts*. It should also be borne in mind that whenever we speak of a watt or a kilowatt, these terms refer to units of *power*; that is, they represent a given expenditure of energy *per second*. Thus, a 40-watt lamp is a lamp that consumes

energy at the rate of 40 watts, which is equivalent to a consumption of energy at the rate of $40 \times 10{,}000{,}000 = 400{,}000{,}000$ ergs per second.

Since one horse power is equivalent to 746 watts, we may write

$$horse\ power = \frac{watts}{746}$$

404. Total Energy Expended by a Current. Since the watt and kilowatt represent the expenditure of energy per second, it follows that if we desire to compute the *total energy* expended by a current in a given time (one hour, say), we must multiply the power (watts or kilowatts) by the time which the current flows. The units thus employed to express the total energy expended are the watt-hour and the kilowatt-hour. A *watt-hour* represents an expenditure of energy at the rate of one watt for one hour, and a *kilowatt-hour* is an expenditure of energy at the rate of one kilowatt for one hour; that is

$$watt\text{-}hours = watts \times hours, \text{ and}$$
$$kilowatt\text{-}hours = watt\text{-}hours/1000$$

Example 1. A current of 0.5 ampere under a pressure of 110 volts flows through an incandescent lamp for 5 hours. (a) Find the power expended in watts. (b) Find the energy expended in watt-hours. (c) Find the energy expended in kilowatt-hours. *Solution*: (a) Watts = EI; hence $110 \times 0.5 = 55$ watts. (b) Watt-hours = watts $\times$ hours $= 55 \times 5 = 275$ watt-hours. (c) 1 kilowatt = 1000 watts; therefore 275 watt-hours = 0.275 kilowatt-hour.

Example 2. Find the energy expended (Ex. 1) in (a) ergs, (b) joules. *Solution*: (a) Since 1 watt = 10,000,000 ergs per second = 1 joule per second, the total expenditure of energy = 10,000,000 $\times$ watts $\times$ time in seconds $= 10{,}000{,}000 \times 55 \times 5 \times 60 \times 60 = 99 \times 10^{11}$ ergs. (b) One joule $= 10^7$ ergs, hence 99×10^{11} ergs $= 99 \times 10^4$ joules.

EXERCISES. **11.** Given a tungsten lamp, across the terminals of which there is impressed an e.m.f. for 110 volts, giving rise to a current of 0.25 ampere, which flows for 4 hours. Find the heat expended in calories.

12. Find the power expended on the lamp in (a) watts, (b) kilowatts.

13. Find the total electrical energy expended during the 4 hours, in (a) watt-hours, (b) kilowatt-hours.

14. Find the total energy expended on the lamp (Ex. 11) in (a) ergs, (b) joules.

15. What will be the cost of running this lamp for 4 hours at the rate at which electrical energy is sold in your town?

405. The Watt-hour Meter. A *watt-hour* meter, Fig. 370, is an instrument for summing up the energy delivered by an electric circuit. By means of a system of dials, similar to those on a gas meter, the total electrical energy delivered to the consumer in a given time is expressed in kilowatt-hours. The consumer thus pays for the electricity at so much per kilowatt-hour.

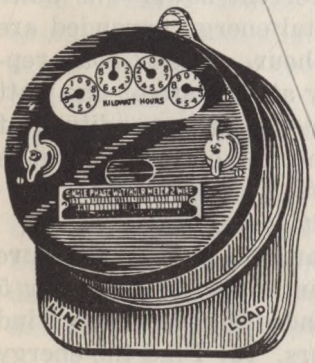


FIG. 370
Watt-hour meter

The mechanism of a watt-hour meter is essentially that of a wattmeter and a small electric motor, the armature of which revolves at a speed which is proportional to the watts used. The rotating part (armature) is geared to a set of recording dials which



FIG. 371.—Recording dials of a watt-hour meter

register the kilowatt-hours in units, tens, hundreds, and thousands respectively, as shown in Fig. 371.

Electric Lighting

406. The Carbon Electric Arc. In the order of the development of electric lighting the arc lamp came first. As early as the days of Sir Humphrey Davy it was known that when the ends of two rods of carbon, suitably connected to the poles of a powerful battery, were brought into contact and then separated a short distance apart, a brilliant light was emitted.



Fig. 372
Enclosed arc

The essential features of an arc lamp, Fig. 372, consist of two sticks of carbon, the ends of which touch. When a current is passed through the line, the carbon points are drawn apart for a short distance, usually by means of an electromagnet connected with the lamp. The current passes across the air gap thus formed and heats the tips of the carbons to incandescence. One carbon, the upper one, in Fig. 373, is called the positive carbon and the other the negative. In the space between the carbons there is formed a volatilized carbon arc which serves as a conductor from one point to the other; it is called the *electric arc* because of its curved shape. The positive carbon is very much the hotter of the two and furnishes the greater part of the light. In the positive carbon there is a depression which is called the *crater*. The temperature of this crater when the lamp is heated is 3500° C, or above, which is the highest temperature attainable by present methods. All known substances are volatilized in the electric arc.

The electric arc will burn under water, but of course

with diminished intensity. The water is decomposed into hydrogen and oxygen.



FIG. 373
Electric arc

Electric arc lights are used now mainly for powerful searchlights. The usual current is from 5 to 10 amperes. In military and naval searchlights, however, currents as high as 100 amperes are used. Some of these searchlights, in which the carbons are cooled with a blast of air, reach a brightness which is estimated at from 200,000 to 300,000 candle power per square inch.

407. The Mercury Arc Lamp. The Cooper-Hewitt *mercury arc lamp* consists of an exhausted glass tube, *AB*, Fig. 374, having at the elevated end *A* the positive electrode which is made of iron, and at the end *B* the negative electrode, consisting of some mercury which is contained in a receptacle or bowl, *B*. In this lamp the incandescent medium is a column of mercury vapor, extending within the tube from *B* to *A*. Under a sufficiently high difference of potential between the electrodes a mercury vapor arc is formed. This arc emits a brilliant and characteristic greenish light, which is almost entirely wanting in red rays.

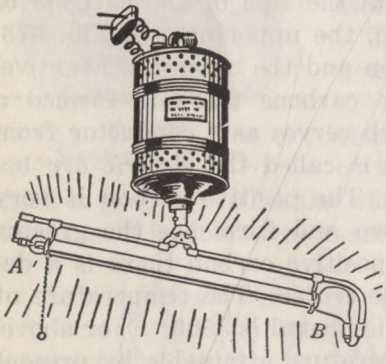


FIG. 374.—Mercury arc lamp

Mercury vapor lamps are extensively used in photography because of the strength of the actinic or chemical active rays, and also for illuminating purposes in certain manufacturing processes.

408. The Incandescent Lamp. At the present time the most important commercial application of the heating effect of a current is in the production of light, 500,000,000 incandescent lamps being in use in the United States alone, which is about half the number used in the entire world. In the incandescent lamp a filament of some not easily fusible substance is enclosed within a glass bulb, from which the air is exhausted. The object of removing the air from the bulb is to prevent the burning out of the filament due to the presence of oxygen.



Edison

The first commercial incandescent lamps brought out contained carbon filaments; these are known as carbon lamps. The filament of the modern incandescent lamp is made of wire drawn from the metal tungsten; these lamps are known as tungsten lamps. The development of this type of incandescent lamp, Fig. 375, was made possible by the discovery that the metal tungsten could be drawn into fine wire, and that this wire could be heated to bright incandescence without reaching the melting point. This important discovery was made by William D. Coolidge in 1911.

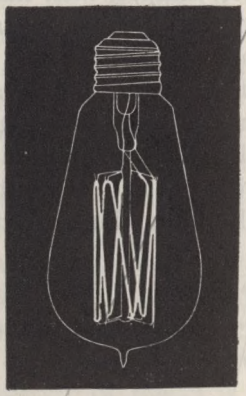


FIG. 375
Tungsten incandescent lamp

Incandescent lamps are in general connected in parallel between 110 volt mains which lead from the dynamo (direct or alternating) or from the low voltage side of a transformer (Art. 454). The method of connecting incandescent lamps in parallel is shown in Fig. 376. Individual lamps of a circuit may be turned on or off by means of a key in the socket; all the lamps of the circuit may be turned on or off by means of a switch S in the line. The resistance of the incandescent lamp is comparatively high, that of the 40-watt tungsten lamp being, when hot, about 250 ohms. When the lamps are connected in parallel, the resistance of the system, however, is very materially reduced. Also, when the lamps are in parallel, one or more may be turned off without seriously disturbing the rest of the circuit.

409. The Gas-filled Tungsten Lamp. In 1879 Thomas Edison produced his first commercially successful carbon-filament vacuum lamp. In 1913 Irving Langmuir invented the modern gas-filled tungsten lamp, which has an efficiency under certain conditions 6 or 7 times as great as that of the old carbon lamp.

A sectional outline of a gas-filled tungsten lamp is shown in Fig. 377. The filament consists of a tungsten wire wound into a very fine spiral coil S which communicates electrically through the conductors C and C' with the external contacts at the base of the bulb. The object

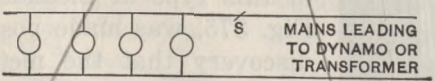


FIG. 376.—Incandescent lamp circuit

of putting the tungsten filament in the form of a fine spiral is to reduce heat radiation. The bulb after being thoroughly evacuated to remove all traces of air and water vapor, is filled with an inert gas, such as nitrogen

or argon, usually a mixture of the two. The object of this inert gas is to exert a pressure upon the tungsten and thus reduce the vaporization of the white hot metal filament. In the presence of an inert gas tungsten may be heated to a very high temperature, thus increasing its illuminating power. Some vaporization of the filament, however, does take place. This metallic vapor is carried upward by the convection currents of the gas and is deposited at the top of the bulb.

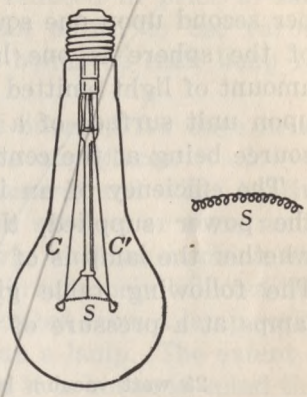


FIG. 377.—Gas filled tungsten lamp

The little point at the bottom of the bulb shown in Fig. 375 is where the tube through which the bulb was evacuated was sealed off. Modern bulbs are rounded at the large end, Fig. 377. Evacuation in such bulbs is effected through a tube at the small end. The sealed end of this tube is covered up by the metal contacts which serve as a connection between the bulb and the lighting circuit.

Most modern electric light bulbs are smooth on the outside and frosted on the inside. Bulbs which are frosted on the outside absorb from 6 to 20 per cent of the light; those which are frosted on the inside by a special process transmit practically all the light emitted by the filament.

410. Electric Lamp Efficiency. The *life* of a lamp is the number of hours it will burn under normal conditions. The life of the modern lamp of the Mazda type is about 1000 hours. The life of a lamp has nothing to do with its efficiency. The *efficiency* of an electric lamp is measured by the number of lumens per watt of power supplied. If

a standard candle is placed at the center of a hollow sphere of radius one foot, the amount of light emitted per second upon one square foot of the interior surface of the sphere is one lumen; that is, a lumen is the amount of light emitted per second by a standard candle upon unit surface of a sphere of unit radius, the light source being at the center of the sphere.

The efficiency of an incandescent lamp depends upon the power supplied, the kind of filament used, and whether the lamp is of the vacuum type or is gas-filled. The following table gives the efficiency of tungsten lamps at a pressure of 115 volts:

25 watt vacuum lamp, 9.5 lumens per watt
50 watt gas-filled lamp, 10.5 lumens per watt
100 watt gas-filled lamp, 13.4 lumens per watt

411. Causes of High Bills for Electricity. Customers pay for electricity on the meter plan at so much per kilowatt-hour. Under ordinary circumstances all electricity consumed, whether for heating, lighting, or power purposes is recorded by the kilowatt-hour meter (Art. 405). When the monthly electricity bills seem unusually high, the meter is the first thing to be suspected; in reality, however, it is usually the last thing to blame. Some of the reasons for abnormally high bills for electric energy are the following:

1. Cloudy or rainy weather, requiring use of light in daylight hours.

2. Additional lamps may have been installed, or small lamps may have been replaced by larger one.

3. Old dim lamps may be in use and in order to get sufficient illumination more of them must be lighted than would be necessary if lamps in good condition were used. A dim lamp takes practically as much current as a new one, and is a very wasteful thing to use. In this connec-

tion it may be well to state that the tungsten lamp has been improved in quality and reduced in price to such an extent that no customer can afford to use carbon lamps, even if he were paid a bonus on each lamp for so doing.

4. Lamps are sometimes left burning for days in attics, closets, and other out of the way places.

5. Electric laundry irons, toasters, and other heating devices may have been placed in service, or some motor driven device such as the electric washing machine may have been installed. It is often erroneously believed that because such devices can be operated from a lamp socket they require no more power than a lamp. The extent of this error may be realized when it is borne in mind that a 6-pound laundry iron takes as much power as 20 tungsten lamps of about 20 candlepower each.

6. Defective wiring may allow current to flow when no lights or other devices are in use.

7. When electric elevators or electrically driven machinery is used and not properly oiled and cared for, excessive friction may result, with a corresponding waste of power and increase in the bill for electric current.

8. An error may be made by the company's meter reader, so that the bill rendered is too high or too low. If it is too high, the bill for the following month will be low by the same amount, if the meter is then read correctly, and consequently the consumer will not lose anything in the long run.

SUMMARY

1. Whenever a current of electricity flows through a conductor, two phenomena are always associated: (a) the current flowing in the conductor, and (b) a magnetic field around the conductor.

2. Rule for determining the direction of the magnetic lines of force around a conductor: Grasp the wire with the right hand,

the thumb in the direction of the current; the fingers indicate the direction of the lines of force.

3. Rule for determining the polarity of a solenoid: Grasp the solenoid with the right hand, the fingers in the direction of the current; the thumb points in the direction of N-pole.

4. Ohm's law: The current in amperes is equal to the e.m.f. or p.d. in volts divided by the resistance in ohms. Ohm's law may be written

$$I = \frac{E}{R}, \text{ or } E = IR$$

5. Laws of resistance: I. The resistance of a conductor is directly proportional to its length.

II. The resistance of a conductor is inversely proportional to its cross sectional area.

III. The resistance of a conductor depends on the material of which it is composed.

IV. The resistance of metallic conductors increases as the temperature increases.

6. The laws of resistance may be written in equational form

$$R = \frac{k \times l}{a} = \frac{k' \times l'}{d^2}$$

7. In metric units, k = ohms per centimeter, l = length in cm, and a = cross sectional area in cm^2.

8. In English units, k' = ohms per mil-foot, l' = length in feet, and d = diameter in thousandths of an inch, and d^2 = circular mils. A mil = 0.001 inch; that is, 1 inch = 1000 mils. A circular mil (C.M.) = d^2 = square of the diameter in thousandths of an inch.

9. Resistance of conductors in series: $R = r_1 + r_2 + r_3 \ldots$

10. Resistance of conductors in parallel: $1/R = 1/r_1 + 1/r_2 + 1/r_3 \ldots$

11. Fall of potential between two points on a conductor carrying a current = $E = IR$, in which E = volts, I = current in amperes, and R = resistance in ohms.

12. Conductors in series: the current strength is everywhere the same throughout the circuit.

13. Conductors in parallel: the currents in the various branches of the circuit depend on the resistances of the respective branches.

14. In a divided circuit between the points B and C, the fall of potential (IR) from B to C is the same over each path, no matter what be the resistance of the conductor.

15. Cells joined in series: $I = \dfrac{nE}{R + nr}$

in which $n =$ number of cells, $E =$ volts, $R =$ resistance of the external circuit in ohms, and $r =$ resistance of the internal circuit.

16. Cells joined in parallel: $I = \dfrac{E}{R + r/n}$

17. Joule's law of electric heating: The heat developed by a current is proportional to the square of the current, the resistance of the conductor, and the time which the current flows. When $I =$ amperes, $R =$ ohms, and $t =$ seconds, the law may be written

$$H = I^2Rt \times 0.24 = \textit{calories}.$$

18. Since $E = IR$, and hence $I^2R = EI$, Joule's law is sometimes written

$$H = EIt \times 0.24 = \textit{calories}.$$

19. B.t.u. = calories/252.

20. The power expended by a current is measured in watts. The watt is defined electrically as volts × amperes, hence

$$\textit{watts} = EI = I^2R$$

21. Horse power = watts/746.

22. The total energy expended by a current in a given time is measured in watt-hours or in kilowatt-hours, that is

$$\textit{watt-hours} = \textit{watts} \times \textit{hours}, \text{ and}$$

$$\textit{kilowatt-hours} = \textit{watt-hours}/1000$$

23. The efficiency of an electric lamp is determined by the number of lumens emitted per candle power. The higher the energy per c.p., the higher the efficiency.

REVIEW EXERCISES

A

1. Discuss the discoveries made by Oersted and Henry.

2. A wire which is stretched in a north-south direction carries a current. (a) It is desired to find the direction of the current, and to this end a magnetic needle is placed below the wire. The N-pole of the needle is deflected to the east. Determine by the right-hand rule the direction of the current in the wire. (b) On another occasion the needle was placed above the wire, the N-pole being again deflected to the east. Find the direction of the current.

3. Wrap a string around a lead pencil after the manner of the solenoid. Consider the current to flow in a given direction in the string, and determine by the right-hand rule the polarity of the point of the pencil.

4. Why is soft iron instead of steel used as the core of electromagnets?

5. State Ohm's law; write the law in the form of an equation.

6. (a) On what four factors does the resistance of a conductor depend? (b) State the laws of resistance in the form of an equation.

7. The resistance of a given piece of wire, length l and diameter d, is 4 ohms. How will the resistance be affected (a) if the length be doubled, d remaining the same? (b) if both length and diameter be doubled?

8. A conductor consists of three wires connected in series, the resistance of the wires being 10, 20, and 30 ohms respectively. What is the resistance of the conductor?

9. A current of 2 amperes flows through the conductor (Ex. 8). Find (a) the fall of potential ($E = RI$) over each of the three wires; (b) the fall of potential over the entire conductor.

10. Two wires of 20 and 30 ohms resistance are connected in parallel between the points B and C. Find the resistance of the two wires thus connected.

11. The total current flowing from B to C (Ex. 10) is 2 amperes. Find (a) the fall of potential from B to C; (b) the fall of potential over each wire; (c) the current in each wire.

12. It is desired to find the resistance of a given conductor, BC, by Ohm's law. A current is passed through the conductor. An ammeter placed in the circuit reads 2 amperes. A voltmeter connected to the terminals B and C indicates a fall of potential of 3 volts. Find the resistance between B and C.

13. A storage battery is being charged. An ammeter placed in the charging line on the positive side of the battery reads 8 amperes. If the ammeter is placed in the charging line on the negative side of the battery, will it read 8 amperes, more than 8 amperes, or less than 8 amperes? Explain.

14. A current of 2 amperes flows through a conductor of resistance 40 ohms for 1.5 hours. Find the power expended in watts; kilowatts.

15. An electric flatiron having a resistance of 50 ohms takes a current of 4 amperes. (a) How many calories of heat are generated in the iron in 3 hours? (b) How many B.t.u.?

16. Compute the cost of operating the iron (Ex. 15) for one hour at the rate per kilowatt hour at which electric energy is sold in your community.

B

17. Explain the operation of the d'Arsonval galvanometer, and tell what advantages it possesses over the tangent galvanometer.

18. Explain the difference of construction between an ammeter and a voltmeter. Why is a shunt used in an ammeter, and why left out in a voltmeter?

19. Find the resistance of 10 m of German silver wire, the diameter of which is 0.4 mm and the value of k is 0.000027 ohm per cm.

20. A current flows through the two wires of problem 10, between the points B and C. The portion of the current flowing through the first wire (20 ohms) is 3 amperes; that through the second, 2 amperes. What is the fall of potential over (a) the first wire? (b) the second wire?

21. A current of 11 amperes flows from B to C over a divided circuit consisting of three wires in parallel. The resistance of the wires is 10, 20, and 30 ohms respectively. Find (a) the total resistance between B and C; (b) the fall of potential between B and C; (c) the current in each wire.

22. Five gravity cells, each having an e.m.f. of 1 volt and an internal resistance of 5 ohms, are joined in series. A wire having a resistance of 25 ohms is connected to the terminals of the battery. Find (a) the resistance of the entire circuit; (b) the current flowing in the wire.

23. Suppose that the 5 cells of problem 22 are connected in parallel. What current will flow through the wire?

24. Find the total energy expended in (a) watt-hours (Ex. 14); (b) joules.

25. The e.m.f. of a battery is equal to the fall of potential around the entire circuit. A wire having a resistance of 2 ohms is connected to the terminals of a cell having an internal resistance of 1 ohm. The current flowing through the circuit is 0.5 ampere. Find (a) the fall of potential over the external circuit; (b) over the internal circuit; (c) around the entire circuit. (d) What is the e.m.f. of the cell?

26. Three dry cells, each having an e.m.f. of 1.5 volts and an internal resistance of 1 ohm, are connected in series to a conductor having a resistance of 7 ohms. Find (a) the resistance of the entire circuit; (b) the current in the conductor; (c) the fall of potential over the conductor; (d) the fall of potential over the internal circuit; (e) the total fall of potential around the

entire circuit. (f) How does this total fall of potential compare with the e.m.f. of the battery?

27. A lead storage battery consisting of 10 cells connected in series shows on being tested a terminal e.m.f. of 16 volts. (a) Does it need to be charged? Explain. (b) Can it be charged by a current which is under a pressure of 15 volts? Why?

28. A street car uses 100 amperes at 600 volts pressure. What power is delivered to it in (a) kilowatts? (b) horse power?

29. How many B.t.u. are given off in an electric oven that takes 10 amperes at 110 volts for 1 hour?

30. An electric heater supplies heat at the rate of 600 B.t.u. an hour. How much power (h.p.) does it require?

31. Read the electric meter dials shown in Fig. 371.

C

32. Find the resistance of 10 ft. of No. 30 iron wire at 20° C. To find k' for iron at 0° C, see Table XVII; to find the diameter of No. 30 wire in mils, see Table V.

33. Given three like cells, e.m.f. of each 1 volt, internal resistance of each 3 ohms. Which is the better method of joining these cells (series or parallel), the external resistance being 3 ohms?

34. The current for an open arc lamp is 10 amperes. The potential at its terminals is 50 volts. (a) What is its resistance? (b) What resistance must be put in series with the lamp if used on a 110-volt circuit?

35. What is the object of using a neutral gas such as nitrogen or argon in the modern gas filled tungsten lamps?

36. Why is the tungsten filament used in the form of a fine spiral?

37. A group of electric lamps in a given circuit are illuminated. When any one of the lamps in the group is unscrewed from its socket, all the lamps go out. How are they connected, in series or in parallel? Explain.

38. Explain why turning off one lamp of the ordinary household lighting circuit does not materially affect the other lamps in the house.

NOTE: For additional problems, see Appendix.

CHAPTER XIII

ELECTROMAGNETIC INDUCTION

Induced Electromotive Force

412. Introductory. Practically all of our commercial electric currents today are obtained from dynamos by the process known as electromagnetic induction. The study of induced electromotive forces and induced currents is therefore of prime importance to the student of electrical phenomena. Oersted discovered that an electric current gives rise to a magnetic field about a conductor, and later on Faraday discovered that a changing magnetic field may give rise to a current. Faraday's famous experiment in electromagnetic induction may be repeated in any laboratory where there is available a coil of insulated wire, a galvanometer, and a magnet. If one end of a magnet (the N-pole, say) be thrust into a coil of insulated wire, the free ends of which are connected to a galvanometer, Fig. 378, the galvanometer needle will be deflected in a given direction; when the magnet is withdrawn from the coil, the needle will be deflected in the opposite direction. If instead of an N-pole, an S-pole be used, the respective deflections of the galvanometer needle will be in the opposite directions to those given by the thrusting in and pulling out of an N-pole. This is known in the literature of physics as *Faraday's experiment in electromagnetic induction.*



Faraday

The current which flows through the coil and galvanometer when the magnet is being thrust in or pulled out is called an *induced current,* and the electromotive force developed by the motion of the magnet is called an *induced electromotive force.* This induced e.m.f., and resulting induced current, is caused by the cutting of the magnetic lines of force across the wires of the coil. The induced e.m.f. manifests itself only when the magnet is in motion, that is, while there occurs a change (increase or decrease) of the lines of force cutting across the coil. *The student should make special note of the fact that an induced current flows only while the magnet is in motion and the coil is closed.* If the coil be open while the magnet is in motion, there will occur in the wire, as before, an induced e.m.f., but no current. Induced currents always occur as a result of induced electromotive forces.

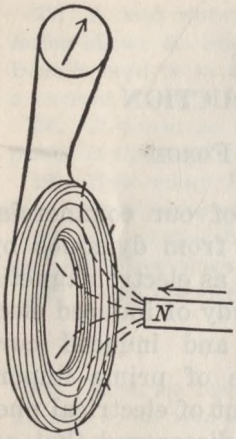


FIG. 378
Electromagnetic induction

413. Relation of the Lines of Force to the Induced E. M. F. We have seen, from the preceding experiment, that when magnetic lines of force cut across a wire, there is set up in the conductor an induced e.m.f. which, if the conductor be closed, gives rise to an induced current. We have also learned that when a current flows through a conductor, lines of magnetic induction are set up in the medium around the wire. That is, currents give rise to magnetic lines of force, and on the other hand, the motion of lines of force may give rise to induced electromotive forces and induced currents. It is important, therefore, to know the relation which exists between the

direction of the motion of the lines of force and the *direction of the induced e.m.f.* There are two rules which are commonly employed to determine this relation, as follows:

1. For the case of a circular conductor the following rule is serviceable. Consider that the N-pole of a magnet is thrust into a coil of wire, Fig. 379, and that the

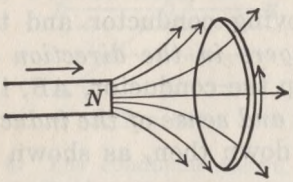


FIG. 379
Indirect induced e.m.f.

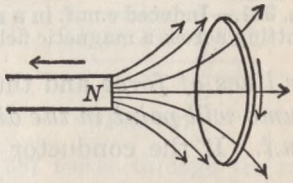


FIG. 380
Direct induced e.m.f.

observer is in such a position as to *look in the direction of the lines of force,* that is, from the S- to the N-pole, at the instant the magnet is thrust into the coil or withdrawn. When the N-pole is thrust into the coil, there is an *increase* of the lines of force threading through the coil which gives rise to an *indirect* induced e.m.f. as shown by the arrow in Fig. 379. When the N-pole is withdrawn, there is a *decrease* of the lines of force which give rise to a *direct* induced e.m.f., Fig. 380. The words indirect and direct as here used refer to the motion of the hands of a watch or clock. *Indirect* means *counter-clockwise; direct, clockwise.* This rule may easily be remembered if it be kept in mind that *i* stands for *increase* and also for *indirect*; likewise *d* stands for *decrease* and also for *direct.*

2. A second rule for determining the direction of the induced e.m.f. is as follows: Suppose that a conductor *AB* moves downward through a magnetic field, Fig. 381, cutting across the lines of force. The direction of the

induced e.m.f. in the conductor is from *A* to *B*. Each line on being cut may be conceived of as tending to wrap itself around the conductor, as shown in Fig. 382. Now if we extend the *right hand* in the magnetic field with the *palm toward* the moving conductor and the *fingers in the direction of the lines of force* and thus grasp the conductor, *AB*, the *thumb will point in the direction and sense of the induced e.m.f.* If the conductor moves down then, as shown in Fig. 382, the induced e.m.f. is "in"; if the conductor moves up, Fig. 383, the induced e.m.f. is "out."

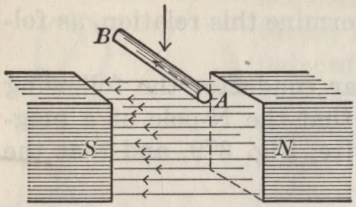


FIG. 381.—Induced e.m.f. in a rod cutting across a magnetic field

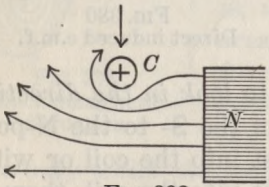


FIG. 382
"In" induced e.m.f.

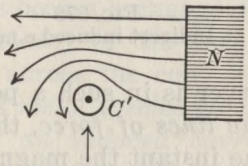


FIG. 383
"Out" induced e.m.f.

EXERCISES

1. A conductor C, Fig. 384, moves through the magnetic field from *C* to *B*. Will the induced e.m.f. be directed toward or from the observer; that is, will it be *in* or *out*?

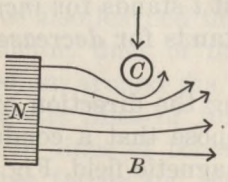


FIG. 384

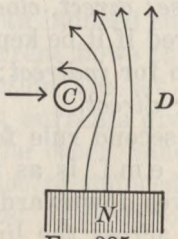


FIG. 385

2. What will be the direction of the induced e.m.f. if the conductor of Fig. 385 move from C to D?

3. Find the direction of the induced e.m.f. when the conductor C, Fig. 386, moves in the direction (a) C to A; (b) C to B; (c) C to F; (d) C to G; (e) D to E.

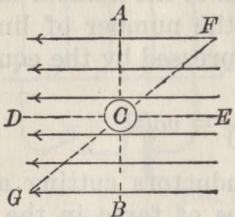


FIG. 386

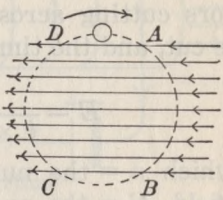


FIG. 387

4. The conductor shown in Fig. 387 moves through the magnetic field in a circular path in a clockwise sense. Find the direction of the induced e.m.f. when this conductor moves from (a) A to B; (b) C to D. (c) At what two places does the induced e.m.f. change its direction?

414. Amount of Induced E. M. F. The *induced electromotive force* in a circuit depends on the *number of lines cut per unit of time.* Now the number of lines cut per second depends, in turn, upon three factors: (a) the number of lines in the field, (b) the number of conductors cutting the lines, and (c) the speed with which the cutting occurs.

In the case of a magnet thrust into a coil, the number of lines cut per second will depend upon the pole strength of the magnet, the rate at which it is moved, and the number of turns of wire in the coil. It is therefore evident that if it is desired to design an apparatus which will give a maximum induced e.m.f., it will be necessary to provide for three things: first, a strong magnetic field; second, the use of as many turns of wire in the coil as will be practicable; and third, as high a rate of cutting of the lines as possible. This is practically the problem which presents itself in the designing of many pieces

of electromagnetic apparatus, such as certain types of dynamos.

When a single conductor cuts lines of force at the rate of 100,000,000 per second, the e.m.f. generated is 1 volt. The relation of the e.m.f. generated to the number of conductors cutting across the field, the number of lines of force cut, and the time may be expressed by the equation

$$E = \frac{n \times N}{t \times 100{,}000{,}000} = volts$$

in which n = the number of conductors cutting across the field, N = the number of lines of force in the field, and t = the time in seconds.

Example. A single straight conductor moves across a magnetic field cutting 50,000 lines of force in 0.1 second. What e.m.f. is developed? *Solution*: E = $(1 \times 50{,}000)$ / $(0.1 \times 100{,}000{,}000)$ = 0.005 volt.

415. Self Induction. Joseph Henry discovered that when a current starts to flow in a coil of wire in the form of a helix, there is induced in the coil an opposing e.m.f. which retards the growth of the current; also, when the circuit is opened and the current in the coil begins to die away, there again appears an induced e.m.f. which tends in this case to prolong the current. He found in fact that the greatest current appeared the instant the circuit is broken. This he called the *extra current of self induction.*

The cause of self induction may be explained by means of Fig. 388. Here we have wrapped around a glass or paper tube a coil of wire connected to the poles of a battery. When the key is open, there is of course no current in the circuit. As soon as the key is closed, however, a current flows through the sections of the coil in a clockwise sense. This causes lines of force to appear, exactly as if the N-pole of a magnet had been thrust into

the coil from *A* to *B*. Now the *increase* of these lines of force in the coil produces an *indirect* (counter-clockwise) e.m.f., which opposes the flow of the current, *so long as*

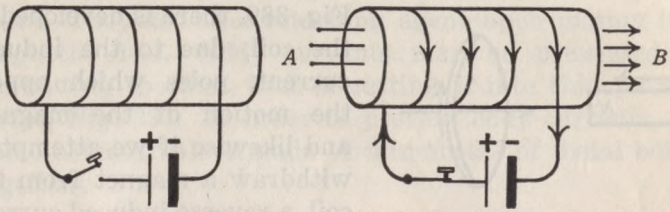


FIG. 388.—Self induction

it is increasing in strength. When the current comes to a steady value, no further change takes place in the number of lines of force in the induced magnetic field, and consequently the opposing e.m.f. ceases. The lines of force, it is true, remain in the field as long as the current continues to flow, but there is no induced e.m.f. so long as there is no change in the number of lines threading through the coil.

Let us consider now what happens when the key is opened. As the current drops, the lines of force in the coil *decrease,* thus producing a *direct* e.m.f. which tends to prolong the current. The effect is the same as if a magnet had been pulled out of the coil.

Thus it appears that when a current flowing through a coil is changing in value, that is, increasing or decreasing, there is always an e.m.f. of self-induction, which tends to oppose the increase of the current the instant the circuit is closed, and which likewise tends to prolong the current after the circuit is broken. The opposing e.m.f. which appears while the current is changing (increasing or decreasing) is the *e.m.f. of self induction.*

416. Lenz's Law. This opposition to change in an electromagnetic system is formulated by Lenz's law,

which may be stated as follows: *When a change takes place in an electromagnetic system, that thing happens which tends to oppose the change.* Thus, by way of illustration, if we attempt to thrust a magnet into a coil, Fig. 389, there is developed in the coil, due to the induced current, poles which oppose the motion of the magnet; and likewise if we attempt to withdraw a magnet from the coil, a reverse induced current is produced which in turn gives rise to a magnetic field opposing the motion of the magnet. This all means that whenever we produce any change in the magnetic system, work has to be done in overcoming the opposing forces. Lenz's law has a very wide application in electromagnetics.

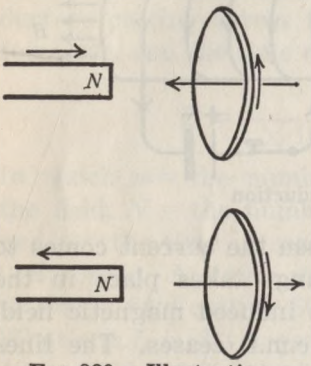


FIG. 389.—Illustrating Lenz's law

417. Eddy Currents. If a magnet be made to rotate about its center above a metal disc, induced currents will be developed in the disc. Suppose that the N-pole of the magnet rotates in a clockwise sense, as shown in Fig. 390. As it moves from left to right, there will be a decrease in the number of lines of force in that part of the disc marked *S*, and an increase in the region *N*. As a consequence, a *direct* induced current having an *S*-pole is developed to the left of the moving magnet, and an indirect current on the right, having an N-pole. Now the *S*-pole pulls back on the *N*-pole of the magnet and the *N*-pole repels it, thus giving rise to a drag which tends to retard the motion of the magnet. If the disc is free to move about its axis, it will, in consequence of the *S*- and *N*-poles developed, rotate with the magnet.

On the other hand, if the disc is rotated the magnet will tend to move with it.

The induced currents which are thus developed in the disc by the changing magnet field are called *eddy currents*. They act as a retarding agent upon motion in a magnetic field. Eddy currents may be prevented by laminating the disc, that is cutting it into thin parallel strips, Fig. 391. In order to prevent eddy currents, dynamo armatures are made of thin sheets of metal bolted together.

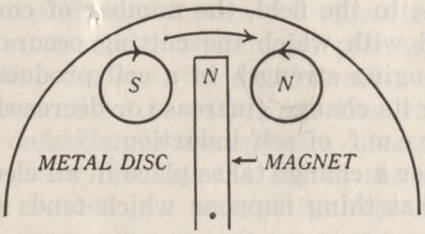


FIG. 390

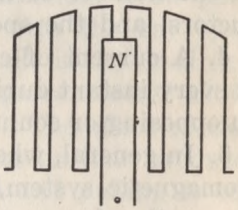


FIG. 391

Eddy currents

418. Fundamental Principles of Electromagnetic Induction. We have learned that when a current flows through a conductor, there is produced in the medium around the conductor a magnetic field; we have also learned that when lines of force thread through a coil or cut across a conductor, there is set up in the conductor an induced e.m.f. If the conductor is in the form of a closed circuit, this induced e.m.f. gives rise to an induced current. Also, whenever a current flowing in a coil changes (increases or decreases), there is set up in the coil an opposing e.m.f. of self induction. The principles involved in these statements of fact are of fundamental importance; they may for convenience be summarized briefly as follows:

1. An increase in the number of lines of force thread-

ing through a coil produces an indirect e.m.f.; a decrease, a direct e.m.f. This induced e.m.f. may be produced either by the motion of the magnet or the coil.

2. Whenever lines of force are cut by a conductor, either by the motion of the conductor or the field, there is produced in the conductor an induced e.m.f., the direction of which may be determined by the right hand rule.

3. The magnitude of the induced e.m.f. is directly proportional to the number of lines per conductor cut per unit of time; the number of lines cut per second depends on the number of lines in the field, the number of conductors, and the speed with which the cutting occurs.

4. A current of changing strength in a coil produces at every instant during its change (increase or decrease) an opposing or counter e.m.f. of self induction.

5. In general, whether a change takes place in an electromagnetic system, that thing happens which tends to oppose the change.

Electromagnetic Appliances

419. Induction by Currents. Currents may be induced by currents. This is not at all surprising since, as we have learned, a coil of wire carrying a current acts like a magnet in that it possesses a magnetic field and polarity. If the coil carrying the current, called the primary, be brought near a second coil, the secondary, an e.m.f., and consequently a current, will be produced in the secondary coil. This may be demon-

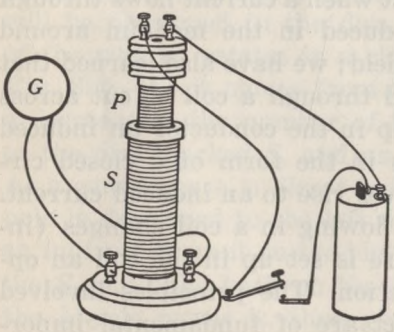


Fig. 392.—Induction by currents

strated with apparatus such as is shown in Fig. 392, in which P is the primary coil and S the secondary, each consisting of coils of insulated wire on hollow, non-conducting cylinders. The primary is connected to a battery and the secondary to a galvanometer. When the key is closed, a magnetic field is produced in P. If the primary now be thrust into the secondary, a momentary deflection of the galvanometer needle will indicate the presence of an induced current in S. When the secondary is withdrawn, the deflection of the galvanometer needle is in the opposite direction. Now place the secondary within the primary, with the key open. Close the key; the galvanometer needle is deflected in a given direction. Open the key; the galvanometer shows a deflection in the opposite direction. The effect of closing and opening the key is the same as that produced by thrusting a magnet into the coil and pulling it out.

Next, place a soft iron core in the primary. Again close the key and then open it. The induced e.m.f. in the secondary is very much increased, as shown by the corresponding increased deflections of the galvanometer needle. The reason for this is that the number of lines of force is increased by the presence of the soft iron core.

420. The Microphone. The *microphone* is a device for producing and detecting slight changes in the strength of an electric current due to corresponding slight changes in resistance. The body of the apparatus, Fig. 393, acts as a sounding board. Fastened to the sounding board by binding posts are two carbon blocks, A and B. Electrical connection between A and B is made by a carbon rod whose tapering ends rest loosely in conical depressions in the blocks, which are placed in circuit with a battery and a telephone. When the circuit is closed, the least motion of the sounding board, as caused by the ticking of a watch, moves the loose carbon pencil, thus varying

the pressure between its ends and the supporting bars. A slight increase of pressure between the carbon contacts reduces the resistance. Vibrations of the sounding board, therefore, cause variations in the pressure at the points of contact of the carbons, as a result of which there are corresponding fluctuations in the current and vibrations of the telephone disc.

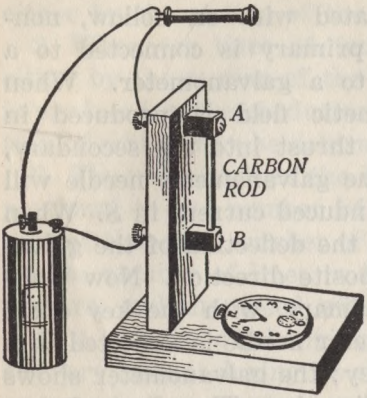


FIG. 393.—Microphone

The varying resistance of carbon under varying pressure makes it a valuable material for use in telephone transmitters, granular carbon being used, however, in place of the carbon rod of the microphone.

421. The Telephone. The electric telephone was invented by Alexander Graham Bell in 1875. In Fig. 394 there is shown the original crude telephone used by Bell in the first successful reproduction of speech electrically.



Alexander Graham Bell

The characteristic features of a simple telephone are shown in Fig. 395. The working parts of a telephone system are the transmitter *T*, the receiver *R*, electromagnet *M* with primary and secondary coils, the local battery, and the line *L*. The operation of the system may be briefly explained thus: When a person speaks into the transmitter *T*, the metallic diaphragm is set in vibration, pressing against the granulated carbon re-

sistance c, and thus producing a variable resistance in the local circuit and hence a variable current. This varying current produces an induced current in the line which affects the magnet of the receiver R', the diaphragm of which is set in vibration. It is the vibration of this diaphragm which produces the characteristic tones of the telephone. When a person speaks into the transmitter T', the same transmission of vibrations occurs with respect to the receiver R.

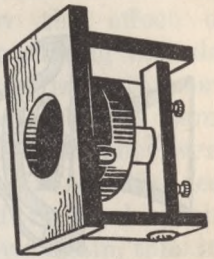


FIG. 394
The first electric telephone, as produced by Alexander Graham Bell

Of course the transmitter and receiver systems in modern telephone circuits are very much more complicated than those shown in Fig. 395; the principle upon which they operate, however, is not essentially different from that just explained. Sectional

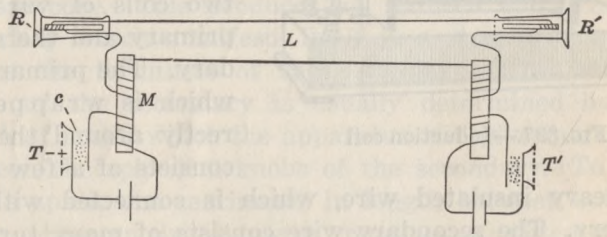


FIG. 395.—Diagram of simple telephone system

diagrams of the modern transmitter and receiver are shown in Fig. 396.

422. The Induction Coil. The *induction coil* is a device for transforming a continuous current of low voltage to an intermittent current of very high potential. Induction coils are extensively used to furnish a spark to

ignite the gas in gas engines, to operate X-ray tubes, and for various other purposes.

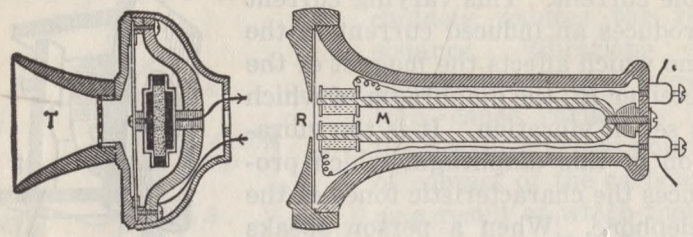


FIG. 396.—Modern transmitter and receiver

The appearance of the ordinary laboratory coil is shown in Fig. 397, a diagram of this coil being given in Fig. 398. The coil is made of a bundle of soft iron wires to prevent eddy currents. Surrounding the core are two coils of wire, the primary and the secondary. The primary coil which is wrapped directly around the core consists of a few turns of heavy insulated wire, which is connected with the battery. The secondary wire consists of many turns of very fine wire, and is not connected in any way with the primary. Underneath the coil is a condenser, which consists of a number of sheets of tin foil carefully insulated from each other.

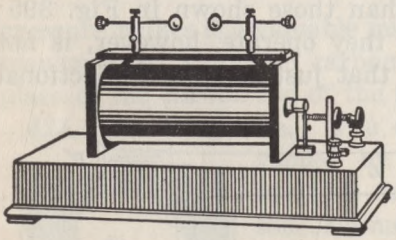


FIG. 397.—Induction coil

The operation of the coil is as follows: When the key is closed, a current flows around the primary circuit, thus magnetizing the core, which in turn attracts the soft iron hammer *a*, which is held against *b* by a light spring.

This breaks the primary circuit exactly as in the case of the electric bell. Now the effect of magnetizing and demagnetizing the core is exactly the same as if a magnet were thrust into the secondary and withdrawn again, thus inducing alternating currents. Since the induced e.m.f. in the secondary depends upon the number of lines of induction cut per second and the number of lines in turn depends upon the number of turns of wire in the secondary, the greater the number of turns in the secondary the greater will be the induced e.m.f. When the coil is in operation, the hammer flies back and forth, making and breaking the primary circuit at *b*, thus magnetizing and demagnetizing the core. This produces an induced e.m.f. in the secondary which manifests itself in a stream of sparks across the terminals of the secondary. The induced e.m.f. of the secondary is usually determined by the length of spark which the apparatus will give in dry air between the spherical knobs of the secondary. To produce a spark one centimeter in length between spheres one centimeter in diameter requires a potential of about 25,000 volts. Induction coils have been made which are capable of giving a spark 50 centimeters in length. To do this would require a difference of potential between the knobs of the secondary of more than a million volts. It must be borne in mind, in this connection, that the induction coil does not enable us to increase the energy furnished by the primary circuit. While the e.m.f. of the secondary circuit is very high, the current is pro-

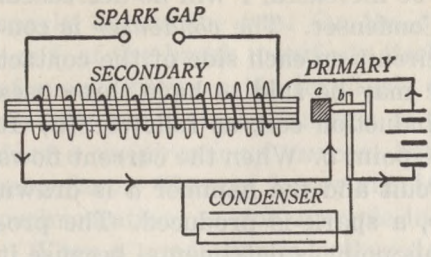


FIG. 398.—Diagram of an induction coil

portionately small. Since the power of a current is equal to EI, we see that if E be increased, I will be decreased.

423. The Use of the Condenser. The *condenser* is connected to the primary circuit on each side of the contact point b. The condenser may be said to have three uses in connection with an induction coil, as follows: (a) It prevents sparking at the point b. When the current flows around the primary circuit and the hammer a is drawn away from the point b, a spark is produced. The production of a spark at this point is detrimental because it tends to prolong the current. When the condenser is in the primary circuit, the energy of the primary current, instead of being expended in the formation of a spark, is expended in charging the condenser, one side becoming positive and the other negative.

(b) Now while the hammer a is drawn to the magnet and the primary circuit is broken, the condenser, which the instant before was charged, now discharges itself back through the battery and around through the primary coil. Now this discharge from the condenser traverses the primary coil in the opposite direction to that of the current which flows from the battery, thus suddenly cutting out the lines of induction in the core. This action of the condenser therefore tends to demagnetize the core more rapidly than would otherwise be possible, thereby *increasing the time rate of the cutting of lines of induction* by the secondary, and consequently *increasing the induced e.m.f.*

(c) The third important function of the condenser grows out of this sudden interruption of the spark at the break. The current at "make" takes a fraction of a second to grow up to its maximum value; while at the "break" by the use of the condenser the cessation is almost instantaneous. Thus the rate of cutting of the magnetic lines of force is much greater at the "break" than at the

"make." The induced e.m.f. at the make lasts longer than at the break, but is feeble and does not suffice to send sparks across the gap. On the other hand, the induced e.m.f. at the break manifests itself by a brilliant torrent of sparks between the terminals of the secondary. Thus we may say that one function of the condenser is to produce a *unidirectional current* between the knobs of the secondary. This is important in certain kinds of experimentation, as, for example in X-ray work.

When a condenser is not used in connection with the induction coil the discharge between the knobs of the secondary is oscillatory.

424. Ignition Systems. An ignition system for a gas engine includes a source of e.m.f. (battery or electric

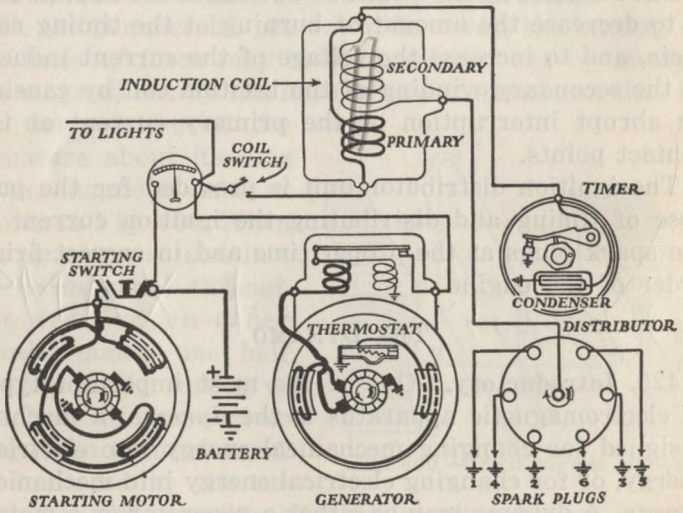


FIG. 399.—Gas engine ignition system

generator), an induction coil, a timer, a condenser, and a distributor. In Fig. 399 there is given a diagram of an automobile ignition system. The induction coil is for the

purpose of converting the low voltage current from a storage battery or generator to a current of very high voltage that will jump the gap in the spark plugs. The system consists essentially of a soft iron core, with a primary winding of a comparatively few turns of copper wire, and a secondary winding of several thousand turns of very fine wire. When the current from the storage battery or generator flows through the primary winding, it magnetizes the core, and when the current is interrupted by the timing contacts, the magnetism dies out. The dying out of the magnetism in the core induces in the secondary winding a high voltage current. This is conducted from the high tension terminal to the distributor head, rotor, and spark plugs.

The function of the condenser, as explained in Art. 423, is to decrease the amount of burning at the timing contacts, and to increase the voltage of the current induced in the secondary winding of the ignition coil by causing an abrupt interruption of the primary current at the contact points.

The ignition distributor unit is provided for the purpose of timing and distributing the ignition current to the spark plugs at the proper time and in correct firing order of the engine.

The Dynamo

425. Introductory. One of the most important types of electromagnetic apparatus is the *dynamo*, a machine designed for changing mechanical energy into electrical energy, or for changing electrical energy into mechanical energy. A dynamo may be either a generator or a motor. When a dynamo is used for the purpose of transforming mechanical energy into electrical energy, it is called a *generator*; when it is used to transform electrical energy into mechanical energy, it is called a *motor*. The funda-

mental principle governing the operation of the dynamo when used as a generator is the production of an induced e.m.f. by the cutting of lines of force, and when used as a motor, the production of mechanical motion due to the reaction of two magnetic fields.

Let us first consider the dynamo as a generator. One of the simplest forms of generators is that exemplified by a coil of insulated wire and a magnet, as shown in Fig. 378, Art. 412. When the magnet is thrust into the coil, lines of force cut across the wires of the armature (coil) and a momentary current flows through the circuit in a given direction. Such a simple generator would give an alternating current, that is, a current flowing at one instant around the circuit in one direction and at the next instant in the opposite direction.

In the commercial generator, however, the cutting of the lines of force may be accomplished not by moving the magnet, but by moving the coil or armature, that is, by the rotation of the armature about its axis in the magnetic field, as shown in Fig. 400. When the armature rotates in a clockwise sense through the arc *abc*, in other words, makes one half revolution, the cutting of the lines of force gives rise to an induced e.m.f. as indicated by the arrows; on the second half of the revolution, that is, *cda*, the cutting of the lines of force is reversed with respect to the conductors, and the induced e.m.f. is consequently in the opposite direction. Thus starting with the armature coil in a vertical position, *ac*, the current, for the

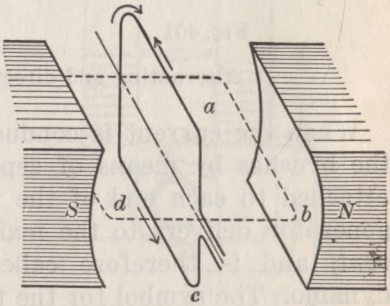


FIG. 400.—Ideal simple dynamo

first half revolution, flows around the coil in one direction, and for the second half revolution, in the opposite direction. *The current in the armature, then, is always an alternating current.*

426. Terms Relating to the Dynamo. The magnets represented by *N* and *S*, Figs. 401 and 402, are called the *field magnets*, and the space between them is called the *magnetic field*. The coil of wire which rotates in this field, is the *armature*. The two metal strips *B* and *B'*, which conduct the current from the armature, are the *brushes*. The conductor through which the current flows is the external circuit, or *main line*.

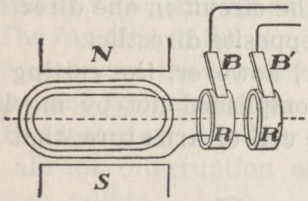


FIG. 401

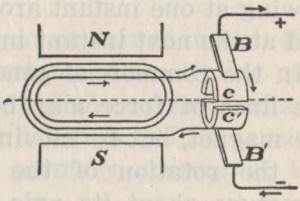


FIG. 402

Alternating and direct current dynamos

When the current is conducted from the armature to the brushes by means of separate rings, Fig. 401, one attached to each end of the wire of the armature, the generator delivers to the main line an *alternating current*, and is therefore called an alternating current dynamo. The symbol for the term *alternating current* is *a.c.*

When the current is conveyed to the brushes by metallic segments, *c* and *c'*, Fig. 402, which device is called a *commutator*, the dynamo is called a *direct current* dynamo. The symbol for the term *direct current* is *d.c.*

427. Kinds of Dynamos. Dynamos may be classified

in various ways. (a) First, with reference to the currents furnished or used, they may be classified as a.c. or d.c.; (b) with reference to the number of field poles employed, as two-pole, four-pole, Fig. 403, or multi-pole, some dynamos of the latter type having as many as 16 poles; and (c) with reference to field winding, as series wound and shunt wound, Fig. 404.

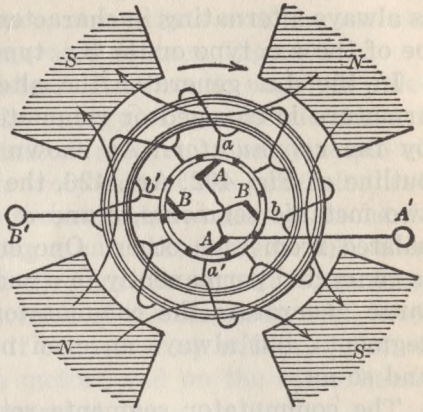


FIG. 403.—Diagram of four-pole dynamo

In the *series* wound dynamo all the current of the main circuit passes around the field circuit. In the *shunt*

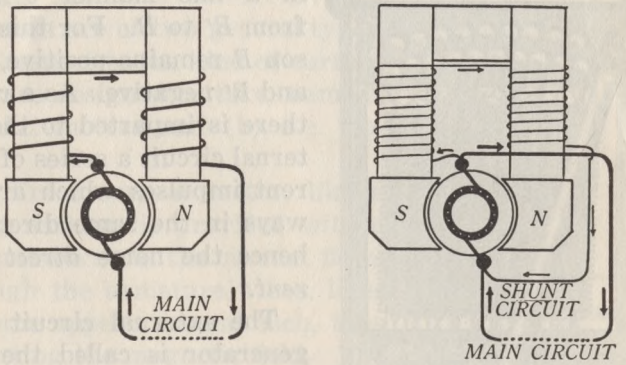


FIG. 404.—Diagrams of series wound and shunt wound dynamos

wound dynamo, on the other hand, only a portion of the current is carried around the field circuit.

428. The D.C. Generator. With respect to the sort

of current furnished, generators are of two kinds, a.c. and d.c. So far as the armature current is concerned, it is always alternating in character, whether the generator be of the a.c. type or the d.c. type.

In the d.c. generator the alternating current in the armature is changed or commuted into a direct current by the *commutator*. As shown in the simple dynamo outline of Fig. 402, Art. 426, the commutator consists of two metallic segments, c and c', which are carefully insulated from each other. One end of the armature coil is connected permanently to c and the other end to c'. In large generators the commutator may consist of many segments, but always an even number, that is, 2, 4, 6, and so on.

The commutator segments rotate with the armature. At the instant the current reverses in the rotating armature and the charges on c and c' change sign, segment c slips from brush B to B', and in a like manner c' slips from B' to B. For this reason B remains positive, say, and B' negative. As a result there is imparted to the external circuit a series of current impulses which are always in the same direction, hence the name *direct current*.

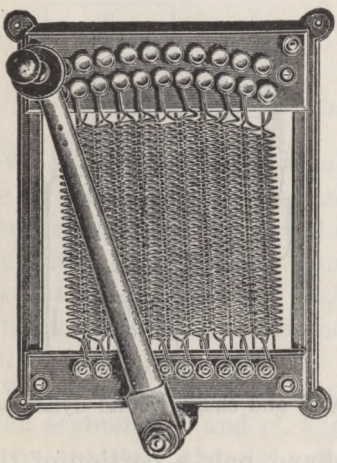


FIG. 405.—Rheostat

The external circuit of a generator is called the line circuit or *line*. The electrical power (EI) expended upon the line is the *load*. The resistance of the line circuit may be changed by changing the load, that is, by switch-

ing lamps on or off, or by means of a rheostat. A *rheostat* is a frame or box containing a series of conductors of relatively high resistance, connections with the various units being made by a movable contact, Fig. 405.

429. The Electric Motor. An *electric motor* is a device for transforming the energy of an electric current into the energy of mechanical motion. The principle of the motor does not differ from that of the generator, except that in one case the armature is rotated by mechanical means, thus generating a current; while in the other case a current is forced through the armature, causing it to rotate, due to a distortion of the magnetic field. Any ordinary d.c. generator, if connected up with a source of e.m.f., will operate as a motor; and on the other hand, any d.c. motor may be operated as a generator. For example, if an electric car be allowed to run down grade, its motor may act as a generator.

Motors are of two general types, with respect to the kind of current they are designed to take, namely, (a) d.c. motors and (b) a.c. motors.

A direct current motor, like a d.c. generator, is provided with a commutator. As the current flows through the armature wires, lines of force appear around each, thus distorting the magnetic field. The principle upon which the motor works may be demonstrated by a device such as shown in Fig. 406. From the contact hooks AB suspend a light U-shaped copper wire so that it swings freely between the poles N and S of two strong magnets. Connect the

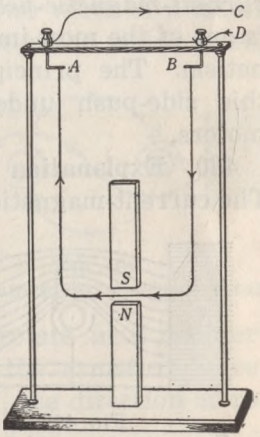


FIG. 406
Magnetic side-push

wires *C* and *D* to the poles of a battery of one or more dry cells, depending on the resistance of the circuit. Place a reversing switch in the battery line *CD*. When the key is closed and the current flows in the suspended loop in the direction shown, the wire between the magnetic poles swings inward, that is, away from the observer; when the current is reversed in the wire, it swings toward the observer.

It thus appears (a) that when a conductor carrying a current is placed in a magnetic field, it tends to move out of the field in a direction at right angles to both the magnetic field and the current, and (b) the direction of motion is reversed when the current is reversed. This *side-push* which is exerted on a conductor carrying a current in a magnetic field is technically known as the *current-magnetic-field force*. The existence of this force is one of the most important facts of electricity and magnetism. The principles involved in the application of this side-push underlie the operation of all electric motors.

430. Explanation of Current-Magnetic-Field Force. The current-magnetic-field force or *side push* is caused by

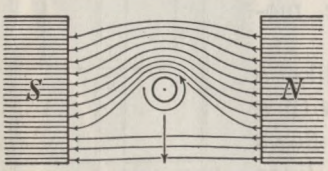


FIG. 407
Current out; motion down

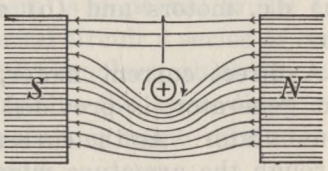


FIG. 408
Current in; motion up

the distortion of the magnetic field by the lines of force around the conductor. Consider Figs. 407 and 408, in which we have represented magnetic fields in the plane of the paper with a conductor in the fields and at right angles to the paper. That is to say, we are looking at

the magnetic fields *side on* and at the conductors *end on*. When there is no current flowing in the conductor, there is of course no distortion of the field.

Let us assume now that there is an *out* current in the conductor. The field is distorted as shown is Fig. 407. Because of the crowded lines of force above it, the conductor tends to move *down*. When the current in the conductor is reversed, that is, when there is an *in* current, Fig. 408, the lines are crowded below the conductor and consequently it tends to move *up*.

In order to find out which way a conductor will move in a magnetic field, all that is necessary is to determine on which side of the conductor there is the strongest magnetic field. *The conductor is always pushed away from the side where the field is strongest.*

In Fig. 409 we have a picture of two wires of a motor armature in the magnetic field between the motor poles. The current is represented as flowing *in* at the right and *out* at the left. The field is distorted downward on one side and upward on the other. Now the tendency of these distorted lines is to straighten, thus causing the armature *AB* to rotate in a counterclockwise sense. If the current in the armature is reversed, that is, *in* at *B* and *out* at *A*, the direction of rotation will be clockwise.

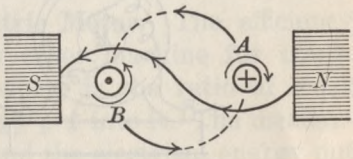


FIG. 409
Diagram of an electric motor

The direction of rotation of a d.c. motor may be reversed by (a) reversing the direction of the current in the armature or (b) by reversing the polarity of the field magnets. If both the armature current and the polarity of the field magnets be reversed simultaneously, no change in the direction of rotation will occur. Most street

car motors are of the series wound d.c. type. Sometimes, however, street cars are operated on a.c. circuits. In such cases the motors are essentially the same as direct-current series-wound motors, for since the current reverses in the field magnets at the same time that it reverses in the armature, it follows that the armature is always impelled to rotate in one direction.

431. Back E.M.F. in a Motor. When the armature of a motor rotates in a magnetic field, as explained in the preceding topic, it cuts some lines of induction and therefore tends to act as a generator. This gives rise to a counter or back e.m.f. in the armature, which opposes the applied e.m.f. at the brushes. *The faster a motor rotates, therefore, the greater is the back e.m.f. developed, and hence the less current it takes.* This fact may be

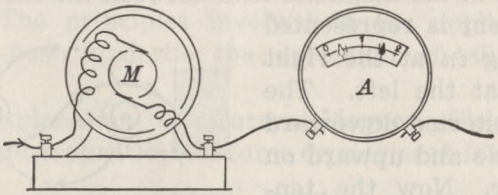


FIG. 410.—Back e.m.f. in a motor

demonstrated by means of a motor, in the circuit of which there is connected an ammeter, Fig. 410. When the current is turned on, the pointer of the ammeter gives a large throw, showing that a large current is flowing through it. As the motor speeds up, however, it develops a back e.m.f. which opposes the applied e.m.f. at the brushes, thus cutting down the current, as indicated by the fall of the pointer.

It is beacuse of the fact that the current through the armature of a motor depends upon its speed that a

"starting box," Fig. 411, is used in connection with motors of any considerable size. A *starting box* is a sort of rheostat so devised as to enable the operator by throwing a switch to reduce the resistance in the circuit and thus gradually increase the e.m.f. as the motor speeds up. If the full current were turned on while the motor is at rest, the chances are that the initial current would be so great that the armature would be burned out. The motorman on the trolley cars uses a form of starting box which he operates in starting his car.

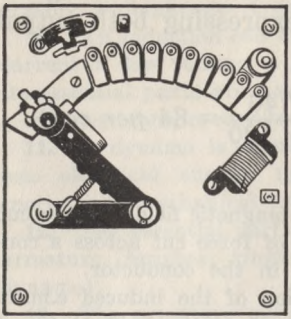


FIG. 411.—Starting box

Direct current motors are used in propelling street cars, in automobile electric starters, and in certain kinds of shop work.

432. Efficiency of the Electric Motor. The efficiency of a motor, like that of any other machine for transforming energy into useful work, is the ratio of work done by the motor to the energy put into it. The mechanical work done by the motor and the electrical energy put into it must of course be measured in units of the same kind. Since *efficiency = output/input,* it is necessary therefore to express both output and input either in foot pounds or in watts, depending on the nature of the units employed.

Output is usually measured in foot pounds or in horse power, which is 33,000 foot pounds per minute. *Input* is usually measured in watts $\times$ time. Since 1 horse power = 746 watts, 1 watt = 33,000/746 = 44.2 foot pounds per minute. Also watts = amperes $\times$ volts $= E\,I$.

Example. An electric motor which takes a current of

14.8 amperes under a pressure of 120 volts develops 2 h.p. What is the efficiency of the motor? *Solution*: Taking the unit of time as the second and expressing both output and input in watts, then

$$efficiency = \frac{output}{input} = \frac{2 \times 746}{14.8 \times 120} = 84 \text{ per cent}$$

SUMMARY

1. Whenever a conductor moves in a magnetic field so as to cut across lines of force or whenever lines of force cut across a conductor, an induced e.m.f. is generated in the conductor.

2. Rules for determining the direction of the induced e.m.f.: I. Circular conductor. Looking in the direction of the lines of force, an increase in the number of lines threading through the coil gives an indirect (counter-clockwise) e.m.f.; a decrease in the number of lines gives a direct (clockwise) e.m.f. II. Conductor moving across the magnetic field. Grasp the conductor with the right hand with the palm toward the moving body and the fingers pointing in the direction of the lines of force; the thumb points in the direction of the induced e.m.f.

3. The magnitude of the induced e.m.f. generated depends on the number of lines cut per second; that is, the induced e.m.f. is determined by the strength of the field, the number of conductors, and the speed with which the cutting occurs.

4. The opposing or counter e.m.f. which appears in a coil when the current is changing (increasing or decreasing) is called the e.m.f. of self induction.

5. Lenz's law: When a change takes place in an electromagnetic system, that thing happens which tends to oppose the change.

6. Eddy currents are currents which are induced in a metal sheet or bar by a moving magnetic field. Eddy currents act as retarding agents on the moving body.

7. A current may be induced in a closed conductor by a magnet in motion or by changing currents in neighboring conductors; that is, currents may be induced by magnets or by other currents.

8. A microphone is a device for producing and detecting slight changes in the strength of a current due to corresponding slight changes in resistance.

9. The essential parts of an electric telephone are a source of e.m.f. (battery or dynamo); an electromagnet with primary

and secondary coils (transformer); a variable carbon resistance controlled by a diaphragm (transmitter); an electromagnet operating a diaphragm (receiver).

10. An induction coil is a device for transforming a continuous current of low voltage to an intermittent current of high voltage. Its essential parts are primary and secondary coils wound around a soft iron core, a current breaker, and a condenser.

11. A dynamo is a machine for changing mechanical energy into electrical energy (generator) or for changing electrical energy into mechanical energy (motor).

12. The essential parts of a dynamo are the field magnets, the armature, brushes, rings (a.c. dynamo) and commutator (d.c. dynamo).

13. The armature current is always alternating in character.

14. A commutator is a device for changing the alternating armature current to a direct current in the external circuit.

15. With respect to field winding, dynamos may be series wound or shunt wound.

16. The current-magnetic-field force or side-push on a conductor in a magnetic field is caused by the distortion of the field by the current in the conductor.

17. The direction of the side push, and consequently the direction of rotation of the armature, may be reversed by reversing the current in the armature or by reversing the polarity of the field magnets. If both the current in the armature and the polarity of the field magnets are reversed simultaneously no change in the direction of rotation occurs.

18. Back e.m.f. in a motor is caused by its tendency to act as a generator. The faster the armature of the motor rotates, the greater is the back e.m.f.

19. Watts = amperes × volts

1 h.p. = 33,000 foot pounds per minute = 746 watts

1 watt = 33,000/746 = 44.2 foot pounds per minute

20. Motor efficiency $= \dfrac{output}{input} = \dfrac{foot\ pounds}{watts \times minutes \times 44.2}$

REVIEW EXERCISES

A

1. Give a brief account of Faraday's experiment in electromagnetic induction.

2. What is the cause of an induced e.m.f. in a conductor?

Under what conditions will such a conductor furnish an induced current?

3. The N-pole of a magnet is thrust into a coil of wire. Is there an induced e.m.f. in the coil when the magnet is (a) in motion? (b) at rest? Explain.

4. Draw a circle on paper or on the blackboard to represent a closed conductor. Assume that the N-pole of a magnet is thrust into the coil thus pictured. (a) Indicate by an arrow the direction of the induced e.m.f. (b) What would be the direction of the induced e.m.f. if an S-pole were thrust in?

5. Assume that the lines of force of a magnetic field pass horizontally through the room from south to north. Assume also that your pencil is a conductor. Hold it in a horizontal position in an east-west direction. What is the direction of the induced e.m.f. when the conductor moves (a) downward? (b) upward? (c) forward or backward in a horizontal direction?

6. (a) What is meant by a direct and an indirect e.m.f. in a circular conductor. (b) Give the rule for determining whether the e.m.f. is direct or indirect when a magnet is thrust into a coil of wire.

7. Give the rule for determining the direction of the e.m.f. in a conductor which is cutting across the lines of force of a magnetic field.

8. (a) Upon what does the magnitude or amount of induced e.m.f. depend? (b) What are the three factors that determine the number of lines of force cut per second?

9. One end of a coil of wire is connected to the pole of a battery. (a) Why does the current not rise to its full value at once when the other end of the coil is connected to the other pole of the battery? (b) When the current comes to its maximum value in the coil, are there any lines of force in the coil? (c) When the circuit is broken, a spark appears. How do you account for the tendency of the current to prolong itself when the circuit is broken?

10. (a) Explain how a current in one coil may be induced by a current in another coil. (b) What is the effect of putting an iron core in the primary coil?

11. Describe the microphone.

12. (a) Name the essential parts of a telephone. (b) What application is made of the microphone principle?

13. In telephoning from one place to another what is transmitted over the wires, sound or electricity? Explain.

14. What is a dynamo? What is the distinction between a generator and a motor?

15. Name the essential parts of a dynamo and give the use of each.

16. Is the armature current in a d.c. generator direct or alternating? Explain.

17. In what respect does an electric motor differ from a generator? Does a d.c. motor have a commutator?

18. (a) What causes the armature of a motor to rotate? (b) Explain the side push on a conductor carrying a current in a magnetic field.

19. Explain how the direction of rotation of the armature of a motor may be reversed.

B

20. (a) How do you account for the fact that there is an induced e.m.f. in the axle of a railway car wheel when the car is in motion? (b) How should a conductor be moved in the earth's magnetic field so as to get the greatest e.m.f. for a given speed?

21. State Lenz's law, and give an illustration of its application.

22. (a) Make a sketch to illustrate the production of eddy currents. (b) Explain how eddy currents retard the motion of a magnet. (c) How may eddy currents be prevented?

23. Make a sketch of an induction coil and explain how it is possible to get a very high voltage at the secondary terminals from a relatively low voltage at the battery terminals.

24. (a) Why is soft iron rather than highly tempered steel used as the core of an induction coil? (b) Why is a bundle of soft iron wires sometimes used?

25. Explain the use of the condenser in an induction coil.

26. Make a sketch of a commutator and explain how it works.

27. Make a sketch to illustrate the distinction between series wound and shunt wound dynamos.

28. Explain why a series wound motor will run when supplied with either a direct or an alternating current.

29. (a) What causes the back e.m.f. in a motor? (b) How is the back e.m.f. affected as the speed of the motor increases?

30. What is a starting box and why is it used in starting heavy d.c. motors?

C

31. What is the effect on the e.m.f. produced by a generator (a) when the strength of the field is increased? (b) when the number of coils of wire in the armature is increased? (c) when the speed of rotation is increased?

32. When the resistance in the external circuit is decreased, what happens to the current in the field coils of (a) a series wound generator? (b) a shunt wound generator? (c) What effect do these changes in the current strength of the field coils produce on the e.m.f. in each case?

33. The resistance of the armature of a d.c. motor is 10 ohms. The terminals of this motor are connected to a 500 volt line. What current would flow through the armature if no starting box were used?

34. When the motor of Ex. 33 is running at full speed, the back e.m.f. is 450 volts. What is the full speed current through the armature?

35. A block of marble weighing 2 tons is hoisted to a height of 40 ft. in 4 minutes by a motor which takes a current of 5 amperes under a pressure of 220 volts. Find the efficiency of the motor.

36. A given motor has a full speed efficiency of 85 per cent. What h.p. will this motor develop when taking a current of 15 amperes under a pressure of 120 volts?

NOTE: For additional problems, see Appendix.

CHAPTER XIV

ALTERNATING CURRENTS

Production of A.C. Electricity

433. Introductory. The subject of alternating currents, which we are about to study, is not only one of the most interesting phases of electricity but it is in certain respects the most important. This is true because of the almost universal use today of this type of current in the everyday affairs of life. For example, of all commercial current furnished for lighting and power purposes, as well as the many other uses to which electricity is put, probably more than 95 per cent is of the a.c. type. In a sense, indeed, we live in an a.c. world. Alternating currents are used to light our streets, offices, and dwellings. In the laundry they are used to heat the electric iron and in the kitchen the electric toaster. They run the sewing machine and the vacuum cleaner. And the a.c. radio set has practically displaced the old battery (d.c.) set.

Direct currents have to be used for electrolytic processes such as silver plating, nickel plating, electrotyping, and copper refining. They are used for charging storage batteries, and for certain manufacturing operations where speed control of the motor is an important factor. Also, on most street car systems, but not all, d.c. power is used.

Aside from the uses mentioned above, and a very few others, alternating currents are coming into almost universal use, and for the following reasons:

(a) Since on a.c. generators slip rings are used, no

commutators being necessary as in the case of d.c. generators, sparking at the brushes is reduced to a minimum. This makes it possible to produce pressures of higher voltage with an a.c. generator than can be safely done with a d.c. generator. Direct current generators may be built with capacities as high as 5000 kilowatts at pressures as high as 500 volts. Alternating current generators, on the other hand, may be built with capacities of 50,000 kilowatts or more, at pressures as high as 15,000 volts. The advantages to be derived from large capacity in a generator unit are that the larger the unit the more efficient it is, and the lower is its cost per kilowatt of capacity.

(b) The second advantage in the use of alternating currents is that by means of a transformer (Art. 454) these currents may be "stepped up" for high tension long distance transmission purposes, and then "stepped down" to such voltages as the customer may desire. Transmission voltage may be as high as 100,000 to 200,000 volts. If pressures higher than these are attained, the current is likely to jump from the wires to the poles or towers or other nearby objects. The usual step down voltage employed for house lighting is 110 volts.

(c) The third advantage is that by the use of alternating currents of high voltage power may be transmitted more economically than by direct currents. The economy in transmission comes from the fact, first, that because of the extremely high voltages employed, smaller wires may be used, and second, there is less heat loss. The higher the voltage the lower may be the current for a given power unit, and the lower the current the less is the heat loss. Indeed, the main advantage in the use of alternating currents is economy of transmission.

434. Difference between Alternating and Direct Currents. The difference between a.c. and d.c. phenomena

may be explained by means of the analogy of the flow of water currents due to the action of force pumps. In Fig. 412 we have a valveless force pump and a water circuit. The piston moves back and forth between a and c, having its greatest speed on each stroke at the middle point b. When the piston moves forward, the water in the circuit flows in one direction and when the piston moves back, the water reverses its direction of flow. That is, the water surges back and forth in the pipe in response to the back and forth motions of the piston. This back and forth flow of the water is similar in every respect to the back and forth flow of an alternating current.

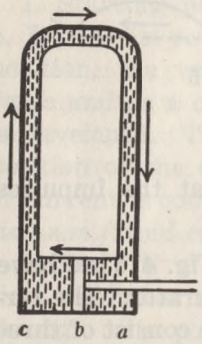


FIG. 412

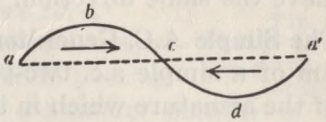


FIG. 413
Illustrating an alternating current

Alternating currents are represented graphically by means of curves such as that shown in Fig. 413. The current in one direction is represented by that part of the curve which is above the axis; that in the opposite direction, by the part of the curve below the axis. The points b and d are the positions of maximum e.m.f. or current. From a to a' on the curve represents one *cycle*. The number of cycles per second is the *frequency*.

A force pump having valves, Fig. 414, furnishes an illustration of direct currents. As the piston moves back and forth, water is forced through the pipes in a series of direct impulses. The valves in the pump correspond to the commutator segments of the direct current generator.

Direct currents are represented graphically as shown in Fig. 415. All the loops are on one side of the axis. A *direct* current is sometimes called a *continuous* current. It is continuous, however, only in the sense that the impulses always have the same direction.

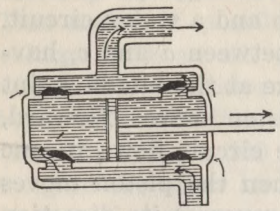


FIG. 414

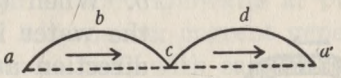


FIG. 415

Illustrating a direct current

435. The Simple A.C. Generator. In Fig. 416 we have a diagram of a simple a.c. two-pole generator. The terminals of the armature which in this case consist of three turns of wire, are connected to the *slip rings* R and R'. Pressing lightly upon the slip rings are the *brushes* B and B', through which the current is conducted to the external circuit. In Fig. 417 we are looking "end on" at one of the conductors of the armature shown in Fig. 416. During one complete revolution the conductor cuts the field twice, once in going from a to b to c during the first half revolution and again in the opposite direction in going from c to d to a during the second half revolution. There is thus generated in the wire an *alternating e.m.f.* which reverses its direction at the points c and a in its passage between successive N- and S-poles of the field magnet. For the first half turn the e.m.f. is directed *in*, that is,

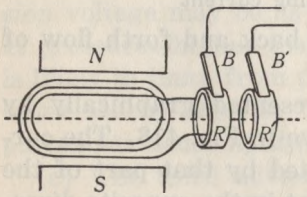


FIG. 416.—Diagram of a simple a.c. generator

away from the observer; second half turn, the e.m.f. is directed *out*.

We desire now to follow this conductor, Fig. 417, step by step throughout one complete revolution, starting at *a*.

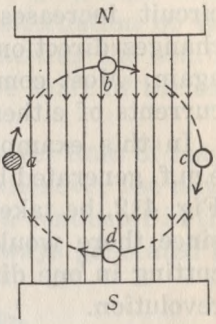


FIG. 417
Change of e.m.f. for one revolution of a conductor

1. Starting position, conductor at *a*. During its revolution through this position, the wire cuts no lines of force and as a consequence no e.m.f. is developed. This is the first dead position of the coil, so far as e.m.f. or current is concerned. *In this position e.m.f. and current are zero.*

2. End of first quarter revolution, conductor at *b*. While passing through this position a maximum number of lines of force are cut per second and hence the greatest direct (in) e.m.f. is generated. *This is the position of maximum e.m.f. and current.*

3. Half turn, conductor at *c*. Here again we have a position in which no lines are cut. This is the second dead point. It is also the point at which the e.m.f. and current reverse; that is, the *in* e.m.f. which existed in the coil during the first half turn now changes to an *out* e.m.f. This change takes place because the conductor, which during the first half revolution cut across the field from left to right, now begins to cut across the field in a reversed sense. *E.m.f. zero; position of change of direction of e.m.f.*

4. End of third quarter, conductor at *d*. *Position of maximum out e.m.f.*

5. Revolution completed, coil back at *a* again. During this fourth quarter turn, the e.m.f. in the conductor decreased until the initial position at *a* is reached when the

e.m.f. again became zero. This completes the *cycle,* that is, *one complete swing or vibration,* (Art. 174) *of the e.m.f.,* during which the current furnished to the external circuit increases from zero to maximum, decreases, changes direction, increases, and then decreases to zero again. Most commercial generators furnish alternating currents of either 25 or 60 cycles per second.

In this example we have considered the alternating e.m.f. generated by a single conductor. If six conductors, Fig. 412, be taken, the e.m.f. will be six times as great, since there would be six wires cutting the field, three cutting in one direction and three in the other for each revolution.

436. Excitation of the Magnetic Field. In general, dynamos (generators and motors) depend for their magnetic induction upon the principle of the electromagnet. Indeed the poles of most dynamos are in reality nothing more than the poles of a big electromagnet. The manner of exciting this magnet depends upon the use to which the dynamo is put. The *magneto* is an a.c. generator, the magnetic field of which is furnished by permanent magnets of the horseshoe type, Fig. 418.

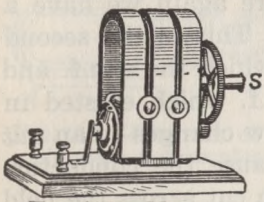


FIG. 418.—Magneto

In the direct current generator there is usually enough residual magnetism in the poles to start at least a small current in the armature upon its rotation. All or part of this current is caused to flow around the field circuit, thus increasing the pole strength. In this way the d.c. generator is said to "build up" its magnetic field.

In alternating current generators this method of exciting the field cannot be directly employed, since the current in the external circuit is an alternating current. In a.c. generators, therefore, the field has to be excited

from some outside direct current source, such as a battery or a small d.c. generator. This is done to keep the polarity of the field magnets from changing sign from N to S and back again during each alternation of the current. The d.c. generator used for exciting the field is sometimes attached directly to the drive shaft of the a.c. generator.

437. Rotating Field Magnets. Whenever lines of magnetic force are cut by a conductor, or whenever the number of lines of force threading through a conductor in the form of a coil is varied, an e.m.f. is always generated in

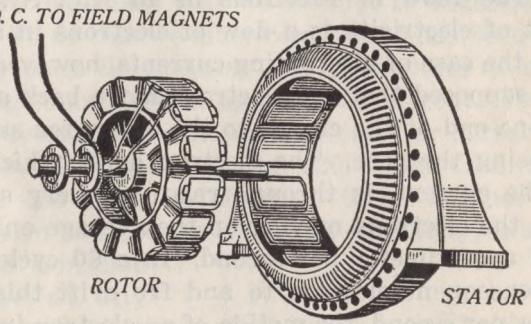


FIG. 419.—Rotating field magneto

the conductor. It is immaterial whether the conductor moves and cuts the lines or whether the lines move so as to be cut by the conductor. In direct current generators, it is always the conductors which move. In other words, direct current generators have rotating armatures. In alternating current generators, especially of the larger capacities, it is usually the field which rotates, although rotating armatures may be used. A stationary armature and a set of rotating magnets are shown in Fig. 419. The wheel with its magnets is called the *rotor;* the stationary armature, the *stator.* The adjacent poles in a rotating field have opposite polarities; that is, if one is an

N-pole, the next is an S-pole, and so on around the entire series. The magnets of the rotating field are excited by a direct current from a d.c. generator. The direct current is conducted into the field through metal rings which are attached to the shaft of the rotor.

The advantage of the rotating field lies in the fact that the armature is stationary and the high voltage current can be taken directly from the armature and delivered to the external circuit without the aid of any sliding contacts. Also, the insulation necessary for high voltage is easier to construct and maintain.

438. The Flow of Electrons in an A.C. System. A current of electricity is a flow of electrons in a conductor. In the case of alternating currents, however, it must not be supposed that the electrons move back and forth from one end of the circuit to the other, for such is far from being the case. The electron drift, which constitutes the current, is through relatively very short distances, the electrons moving on the average only a fraction of a centimeter per second. In a 60 cycle current, with consequent electron to and fro drift this number of times per second, the motion of an electron in a copper wire is perhaps 0.0001 centimeter. In a wire carrying a current of strength sufficiently great enough to cause it to become red hot, the electron motion may be 0.01 centimeter.

Although these distances are extremely small, there is plenty of chance for collisions between the electrons and the molecules of the metal. These collisions produce heat in the conductor. It is estimated that in traversing a distance of 0.0001 centimeter in a copper wire an electron passes about 10,000 molecules of copper.

Voltage and Current Phases

439. Introductory. In driving about the country one

frequently notices that in electric transmission systems there are three high tension wires stretched from pole to pole or from tower to tower. Sometimes a fourth wire is attached to the top of the poles. This is supposed to serve as protection from lightning. The question at once may arise as to why *three* wires are used. One would naturally think that there would be two or four or six wires. If inquiry be made of an electrical engineer, he may offer the information that the three wires are those of a *three-phase* generator. In the study of alternating currents the term "phase" occurs many times. It is therefore important that we attempt to understand its meaning. To do this we must recall what has been stated (Art. 434) as to the meaning and significance of a *cycle,* as applying to an alternating e.m.f. or current. The rise, fall, and reversal of an alternating e.m.f or current may be represented by a curve, Fig. 420, which from a to a' represents one cycle and from a' to a'' another cycle. Now the term *phase,* as employed in electricity, is used to denote the condition of change at any point in a cycle. In engineering practice, phase is measured in electrical degrees. One *electrical degree* is 1/360 of a cycle. For example, at the point a, Fig. 420, where the alternating current is just beginning its positive impulse, the phase is zero; at b, where the current has reached its greatest positive value, the phase is 90°; at c it is 180°; at d, 270°; and at a' where the cycle is completed, it is of course 360°. At this point the cycle begins to repeat its various phases.

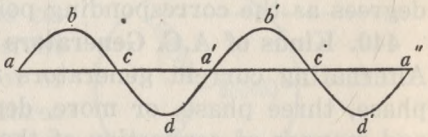


FIG. 420.—Single phase current curve

In the above statement of phase relations, reference in each case is made to electrical degrees. The student should

not confuse electrical degrees with mechanical degrees. An electrical degree is 1/360 of a *cycle*; a mechanical degree is 1/360 of a *circle*. Since one cycle is completed for every passage of an armature coil through a magnetic field generated by consecutive N- and S-poles, it follows that for one revolution of the armature through 360 mechanical degrees the number of cycles developed depends upon the number of coils in the armature and the number of pairs of poles in the field.

Two alternating electromotive forces or currents may be in phase or out of phase. The meaning of the terms *in phase* and *out of phase* may be illustrated by two persons having equal strides and walking side by side. Suppose they start at the same instant, each with the left foot forward, and keep step. They are stepping *in* phase. If, on the other hand, one person starts a little before or a little after the other, they are said to be out of step, or, to use a more technical term, they are *out of phase*. Two electromotive forces or currents are in phase when any point on one curve has the same number of electrical degrees as the corresponding point on the other curve.

440. Kinds of A.C. Generators with Respect to Phase. Alternating current generators may be one phase, two phase, three phase, or more, depending on the number and manner of connection of the coils of the armatures, and the number of poles in the field. We shall consider briefly and in order the one, two, and three phase types of alternators, using the simplest possible generator diagrams. One, two, and three phase generators are represented conventionally by the number and position of the armature coils, the field poles being omitted.

Single phase generator. A single phase generator may be represented by a single coil rotating in a magnetic field. The current, Fig. 420, Art. 439, consists of a single series of cycles.

Two phase generator. In the two phase generator the armature consists of two coils which are set at right

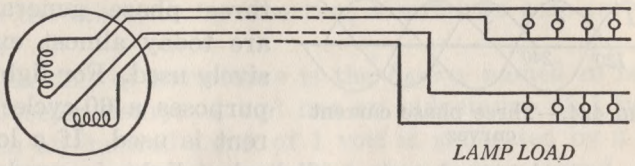


FIG. 421.—Diagram of a two phase generator

angles to each other, Fig. 421. If each coil is connected to a separate pair of slip rings, the generator will deliver two separate and distinct currents. These currents differ in phase by 90°, Fig. 422.

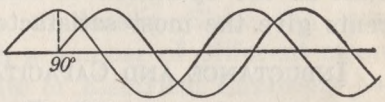


FIG. 422.—Two phase current curves

Three phase generator. The armature of a simple three phase generator has three coils symmetrically placed 120° apart, Fig. 423. If the three coils are connected respectively to three different sets of rings, a generator of this type will deliver three separate currents. This, however, would require six line wires, which for long distance transmission would involve an unnecessary expense because of the large amount of wire that would have to be used. In order to avoid the use of six wires in three phase systems, armature coils have been devised so that only three wires are necessary. During each series of cycles one wire serves successively as a return circuit for each of the other two.

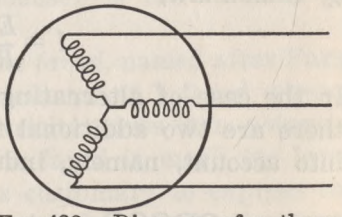


FIG. 423.—Diagram of a three phase generator

Three phase generators give three current curves differing in phase from each other by 120°, Fig. 424. In commercial practice three phase generators are today almost exclusively used. For lighting purposes a 60-cycle current is used. If a lower frequency is employed, a flickering light is produced which is hard on the eyes. For power purposes, however, such as the operation of electric motors, 25-cycle currents give the most satisfactory results.

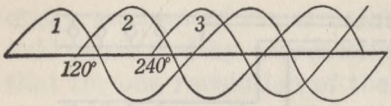


FIG. 424.—Three phase current curves

INDUCTANCE AND CAPACITANCE IN A.C. CIRCUITS

441. Introductory. In our study of continuous currents, as for example, those furnished by batteries and d.c. generators, we had to deal with only three factors, namely, electromotive force E, current I, and resistance R, the relation between these quantities being expressed by Ohm's law,

$$I = \frac{E}{R}, \text{ or } E = IR$$

In the case of alternating currents, on the other hand, there are two additional factors that have to be taken into account, namely, inductance L and capacitance C.

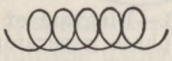


FIG. 425

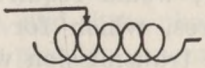


FIG. 426

Fixed and variable inductance

Inductance L is the ratio of the e.m.f. of self induction (Art. 415) in a coil to the rate of change of the current which is flowing through the turns of wire of which the coil is composed. A fixed inductance is, in modern radio practice, represented by a spiral coil, Fig. 425, and

a variable inductance as shown in Fig. 426. The inductance of a given system may be varied by a sliding contact on one of the coils, by thrusting one of the coils into the other, or by inserting a soft iron core into the coil.

The unit of inductance is the *henry,* named in honor of Joseph Henry. A coil has an inductance of 1 henry when an induced e.m.f. of 1 volt is generated by a current changing at the rate of 1 ampere per second.

Capacitance C. In Chapter X, in accordance with long established custom, we used the word *capacity* as applying to a conductor or a condenser. Since the term capacity may be used with a number of different meanings, the American Institute of Electrical Engineers has recommended that when used in an electrical sense, the word *capacitance* be substituted for *capacity.* This recommendation has been quite generally adopted by engineers. The two terms (capacitance and capacity) may be and are used interchangeably, however.

The capacitance of a condenser is the ratio of its charge to its potential.

The unit of capacitance is the *farad,* named after Faraday. A condenser has a capacitance of 1 farad when a charge of 1 coulomb of electricity gives it a potential of 1 volt. Since, however, the farad is much too large a unit for ordinary use, it is customary to express the capacitance of a condenser in microfarads. A *microfarad* is a millionth of a farad.

442. Capacitance of Condensers in Series and in Parallel. Because of the extensive use of condensers in modern inductive systems, telephone and radio apparatus for example, it may be well to consider here the capacitance effects of connecting condensers in series and in parallel.

(a) *Series.* Let us take three condensers having capacities of C_1, C_2 and C_3 microfarads respectively. First,

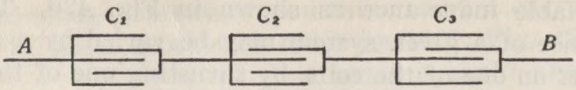


FIG. 427.—Condensers in series

connect the three condensers in series, Fig. 427. The capacitance C of the combination is given by the equation

$$\frac{1}{C} = \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3}$$

The student will note that this is just the opposite effect to that obtained when the resistances R_1, R_2, R_3 are connected in series (Art. 389).

Example 1. Given three condensers having capacities of 0.2, 0.4, and 0.6 microfarad respectively, to find the capacitance of the system when connected in series.

$$\textit{Solution}: \frac{1}{C} = \frac{1}{0.2} + \frac{1}{0.4} + \frac{1}{0.6}, \text{ hence}$$

$$C = \frac{1.2}{11} = 0.1 \text{ microfarad}$$

This means that a number of condensers joined in series may have a capacitance less than that of any one of the series.

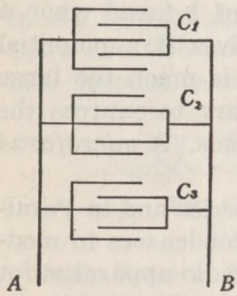


FIG. 428
Condensers in parallel

(b) *Parallel.* Condensers may be joined in parallel as shown in Fig. 428. In this case the capacitance C of the system is equal to the sum of the capacities of the various condensers, as given by the equation

$$C = C_1 + C_2 + C_3$$

Example 2. Find the capacitance of the three condensers of Ex. 1, when joined in parallel.

$$\textit{Solution}: C = C_1 + C_2 + C_3$$
$$= 0.2 + 0.4 + 0.6 = 1.2 \text{ microfarads.}$$

443. Effects of Inductance in an A.C. Circuit. When a coil of wire is placed in an a.c. circuit, two very important effects may result. (a) First, the a.c. current is choked down by the self inductance of the coil. This choking effect may be illustrated by the following experiment: Place a coil having quite a number of turns of rather heavy insulated copper wire in series with a d.c. ammeter. Let the resistance of the coil be, say, 22 ohms. The resistance of the ammeter is negligible. Suppose now that the coil with the ammeter in series be connected to the terminals of a 110-volt d.c. generator. The current through the coil, as measured by the ammeter, is

$$I = \frac{110}{22} = 5 \text{ amperes}$$

Now put in place of the d.c. ammeter an a.c. ammeter and connect the coil and ammeter terminals to a 110-volt a.c. circuit. The current as recorded by the a.c. ammeter is very small as compared with that obtained from the direct current, possibly less than one ampere, depending upon the number of turns of wire in the coil, the length of the coil, and its cross sectional area. This is an illustration of the choking effect of inductance. A coil of wire in an a.c. circuit is sometimes called a *choke coil.* It is important to note that *Ohm's law does not apply to an a.c. circuit in which there is inductance.*

(b) The second effect of inductance in an a.c. circuit is that inductance puts the e.m.f. and current out of phase; the current *lags behind* the e.m.f. To put it another way: in any given cycle, the current comes to its maximum value at a point which is at a greater number of electrical degrees on the axis than

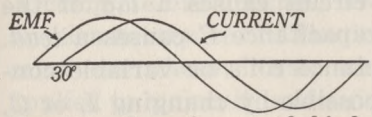


FIG. 429.—Lag of current behind e.m.f.

that of the position of maximum value of the e.m.f. Lag is measured in electrical degrees. In Fig. 429 the current is pictured as lagging behind (to the right of) the e.m.f. by 30°.

The angle of lag is of great significance in the computation of power in an a.c. circuit.

444. The Effect of Capacitance in an A.C. Circuit. If a circuit containing an electric lamp and a condenser, Fig. 430, be connected to the terminals of a 110-volt d.c. line, no effect will be observed. The lamp does not light up, because no current passes through the condenser.

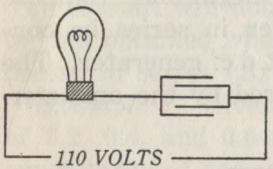


FIG. 430.—Effect of a condenser in a circuit

Now connect the circuit to a 110 a.c. line. The lamp at once lights up. A condenser in an a.c. circuit evidently does not prevent the current from surging back and forth. While no current passes *through* the condenser, the electricity does flow into it and out of it once each cycle. This means that in a 60-cycle circuit the condenser is charged and discharged 60 times a second.

A condenser in a circuit, as just described, throws the current out of phase with the e.m.f., but in this case the current *leads* the e.m.f.; that is, in any given cycle it comes to its maximum value a certain number of electrical degrees before the e.m.f.

Inductance L in an a.c. circuit causes a *lag* of the current behind the e.m.f.; capacitance C causes a *lead*. By means of variable inductance coils or variable condensers, or both, it is thus possible by changing L or C, or both, to bring two alternating electrical systems into phase.

445. Power in an A.C. Circuit. In general, neither Ohm's law nor the power equation, $P = EI$, apply to

alternating currents. The exception is the case of a conductor containing resistance only. For example, if an alternating e.m.f. be applied to a wire in which there is neither coil nor condenser, we may write for current and power the equations which were used in direct currents, namely,

$$amperes = \frac{volts}{ohms}, \text{ and}$$
$$watts = volts \times amperes$$

If, however, there is inductance or capacitance in the circuit, neither of these equations holds true. For Ohm's law we have to substitute another equation, as follows:

$$amperes = \frac{electromotive\ force}{impedance}$$

in which the term *impedance* includes the resistance R of the conductor, the inductance L, the capacitance C, and the number of cycles per second. And for power we have

$$watts = volts \times amperes \times power\ factor$$

The *power factor* is a quantity which depends upon the angle of lag or lead, and must therefore be determined for each individual circuit.

The student will not be able to get from these brief statements regarding *impedance* and *power factor* much of an understanding of the full meaning of these terms. A complete explanation, however, is too complicated for an elementary text. The important facts to be kept in mind are these: first, when an alternating e.m.f. is applied to a circuit containing inductance or capacitance, Ohm's law does not apply; and second, the equation for power, as applying to a.c. circuits with inductance or capacitance, includes a factor which is dependent on the angle of lag or lead.

A.C. Motors

446. Kinds of A.C. Motors. Every motor consists of two essential parts, a stationary part called the *stator* and a rotating part called the *rotor*. There are various kinds of a.c. motors, depending on (a) the stator windings, (b) the kind of current used, that is, single phase, two phase, or three phase, (c) the construction of the rotor, (d) the method of starting, and so on. We shall discuss in detail only two general types, known respectively as the induction motor and the synchronous motor. The principles involved in the construction and operation of these two types are included in one way or another in practically all the other kinds.

447. The Induction Motor. The working of an *induction motor* depends upon the possibility of producing a rotating magnetic field by means of alternating currents. There are single phase, two phase, and three phase in-

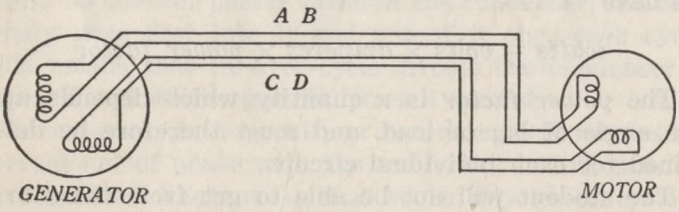


FIG. 431.—Two phase generator and two phase motor

duction motors. The simplest type, so far as explanation is concerned, is that of the two phase motor. A two phase induction motor is supplied with current from a two phase generator, Fig. 431. The stator magnet consists of a ring having four poles, Fig. 432. In considering how the rotating field is produced, it is necessary to recall that when a two phase current is at its maximum in line *AB*, it is zero in line *CD*. In this con-

dition the poles *a* and *b*, Fig. 432, are magnetized and poles *c* and *d* are dead. At this stage, *a* is an N-pole and *b* an S-pole. At the end of a quarter turn of the armature in the generator, current *CD* is a maximum and *AB* is zero, and as a result the poles *ab* swing around to *cd*. At the half turn the currents in the lines reverse, and *b* becomes an N-pole and *a* an S-pole. At the three quarter phase the N-pole moves around to *c*, completing its revolution back to *a* at the end of the cycle. Thus we see that a two phase current gives rise to a *rotating field* in the stator of the motor, the N- and S-poles chasing each other around as the alternations in the lines occur.

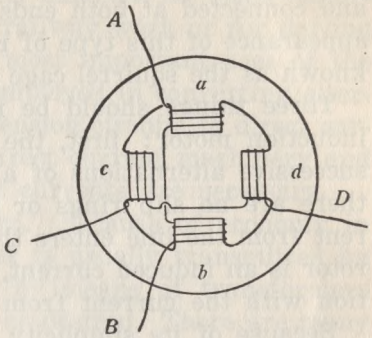


FIG. 432.—Stator magnet, rotating field

If now a permanent magnet, such as a magnetic needle, be pivoted at its middle point and placed with its axis of rotation at or near the center of the rotating field, the poles of the magnet will follow the poles of the field around the field. The rotating magnet corresponds to the rotor of the induction motor. In actual practice the rotor consists of an iron

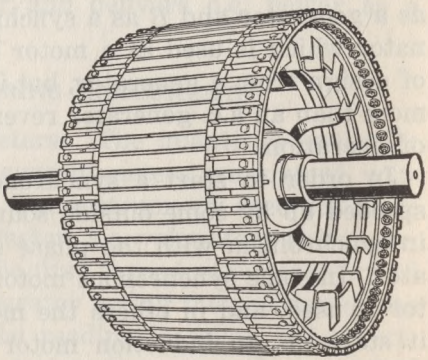


FIG. 433.—Squirrel cage rotor of an induction motor

core with copper bars placed in slots around the outside and connected at both ends to copper rings. From the appearance of this type of rotor to a squirrel cage, it is known as the squirrel cage rotor, Fig. 433.

Three things should be noted with reference to the induction motor: first, the rotating field is due to the successive alternations of a two phase current; second, there are no slip rings or brushes; and third, no current from the line enters the rotor. The current in the rotor is an induced current, but it has no direct connection with the current from the generator.

Because of its simplicity of construction and ease in starting, the induction motor has a very extensive use in factories, and in household motor appliances.

448. The Synchronous Motor. A *synchronous motor* is essentially an a.c. generator operating as a motor. If two a.c. generators, A and B, are connected in parallel, and if while both generators are running B is cut off from its driving power (steam engine or water turbine), it will not stop as might naturally be expected but will continue to run so long as A runs. In this case, A acts as a generator and B as a synchronous motor. An alternator which is used as a motor differs in some features of design from a generator, but in principle it is nothing more than an a.c. generator reversing its normal manner of operation.

In order to start a synchronous motor it has to be speeded up by some outside source until it is in step or in synchronism with the phase conditions of the generator. In some synchronous motors an extra starting motor is used, and in others the motor is so designed that it starts as an induction motor and continues as such until it comes into phase with the generator, after which it runs as a synchronous motor.

The synchronous motor is of use only where constant

speed is possible; it is entirely unsuited for loads requiring variable speed, such as railway work or for driving machine tools. One of the most important uses of the synchronous motor is that employed in converting alternating currents from high tension circuits to direct currents for the operation of direct current machinery and other purposes where direct currents are necessary.

449. The Rotary Converter. Because of economy in transmission electrical power is usually transmitted by alternating currents, which by means of transformers may be stepped up or down as desired. There are many cases in which the power is desired in the form of a direct current, as for example in most electric railway and trolley systems, and in some cities power is distributed to customers as d.c. power. In such cases it is necessary to change the high tension a.c. power to low tension d.c. power. For this purpose the *rotary converter* is largely used. A rotary converter is a synchronous motor, to the shaft of which there is attached a d.c. generator. The converter, therefore, receives a.c. power as a synchronous motor and delivers d.c. power as a direct current generator.

A.C. Measuring Instruments

450. Kinds of A.C. Meters. The quantities generally measured by alternating current meters are current, voltage, power, and energy. The instruments used for these purposes are the a.c. ammeter, a.c. voltmeter, wattmeter, and watt-hour meter. The first three instruments are of the indicator type, the operator noting the position of the pointer at the instant the reading is taken. The watt-hour meter is an energy meter; it sums up by means of a suitable dial system the total amount of electrical energy delivered in a given time.

In addition to the four classes of meters mentioned

above, there are several others which find application in engineering and testing practice. Among these are the *frequency meter* and the *power factor meter.* A frequency meter gives the number of cycles per second. The power factor depends upon the angle of lag or lead between the current phase and the voltage phase. The power factor meter measures this factor, which is necessary to know in order to compute power by means of the equation

$$watts = volts \times amperes \times power\ factor$$

451. A.C. Ammeters and Voltmeters. In general appearance the standard a.c. ammeter, Fig. 434, and a.c. voltmeter is very much like that of the d.c. ammeter and d.c. voltmeter. In the actual construction of their working parts, however, a.c. instruments are quite different from the corresponding d.c. instruments. As explained in Arts. 379 and 380, d.c. ammeters and voltmeters are of the d'Arsonval type, the essential part in each case being a small movable coil between the poles of a permanent magnet. The deflection of the pointer in such an instrument reverses when the current reverses. Since an alternating current is continually reversing, an instrument of the d'Arsonval type would not do at all for an a.c. circuit. The pointer would tend to move back and forth with each reversal. The alternations come so rapidly, however, that what actually would happen is that the pointer would remain at or near

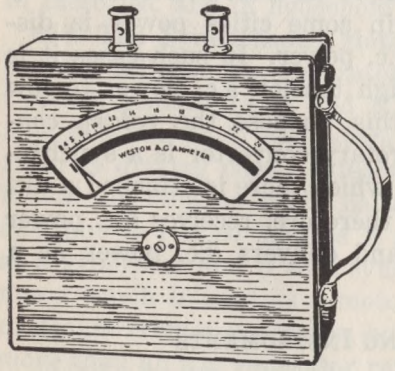


FIG. 434.—A.C. ammeter

zero in a more or less vibratory condition. Some different type of construction must therefore be adopted. If the alternating current meter is to operate by means of a moving coil placed in a magnetic field, it is evident that the magnetic field in which the coil is lying must reverse at the same instant that the current in the coil reverses. In a case of this sort, the coil will continually be deflected in the same direction. The magnetic field, then, in which the moving coil turns, must be generated by the current flowing through the coil. Permanent magnets are therefore of no avail; light coils have to be substituted. From this it follows at once that the weight of the a.c. voltmeter or ammeter may be much less than that of a corresponding d.c. meter, and also that the a.c. meter will use more power than the d.c. type.

When an alternating current or e.m.f. passes through each half cycle, it changes from zero to a maximum value and to zero again, when reversal takes place. In each half cycle there is a *maximum value,* and the average of all the changing values is the *average value.* Now it so happens that a.c. ammeters and voltmeters are calibrated to give neither of these values, but record what are called *effective values.* What is meant by an *effective current,* for example, may be made clear by a consideration of the a.c. ammeter.

An alternating current really has no unit of its own, so we measure it in terms of the direct current unit, the ampere. To have the alternating current ampere exactly equal the direct current ampere, it must produce the same effect as the direct current ampere. Now an ampere is defined as that steady current which will deposit 0.00118 gram of silver from a standard solution in one second. But an alternating current is not a steady current and neither will it deposit any silver from a solution, since whatever it deposits during one half a cycle, it

takes off the next half. Accordingly, in order to compare the alternating with the direct current, we must use some other property which both kinds of current possess. The most natural property is the *heating effect* of each. Therefore an alternating current is said to be equivalent to a direct current when it produces the same heating effect, under exactly similar conditions. This value is called the effective value of an alternating current.

To summarize then, an *effective* ampere, as recorded by an a.c. ammeter, has the same heating effect when flowing through a given wire for a given time as a direct current ampere as measured by a d.c. ammeter. This effective current is less than the maximum current for any given half cycle and somewhat greater than the average current. It may be shown mathematically that the *effective current* is 0.707 times the maximum current.

Likewise a.c. voltmeters are calibrated to give readings in *effective volts*. An effective volt is 0.707 times the maximum voltage attained during any half cycle.

The scale parts on a.c. ammeters and voltmeters are not evenly spaced as they are on the corresponding d.c. instrument. In a.c. ammeters and voltmeters the scale divisions for the lower values are very close together, getting farther and farther apart as one goes toward the greater values.

452. The Wattmeter. A *wattmeter* is an instrument for measuring electric power in watts. The commercial wattmeter is somewhat similar in form to the ammeter and voltmeter, except that the wattmeter has four binding posts, two current terminals, and two voltage terminals, Fig. 435.

The essential parts of a wattmeter are two coils, one suspended within and at right angles to the other. The heavy coil is the *current coil;* it is connected in series

with the power line. The lighter coil is the *voltage coil;* it is connected as a shunt across the power circuit. The magnetic field in the current coil is proportional to the current in the line, and the magnetic field in the voltage coil is proportional to the potential across the terminals *BC*. The like magnetic poles induced in the two coils repel each other, thus causing the voltage coil to rotate on its vertical axis. This rotation moves the pointer over the scale.

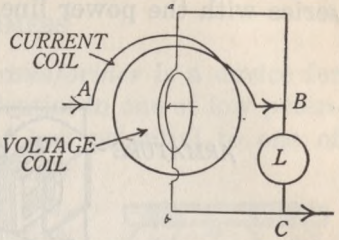


FIG. 435.—Diagram of a wattmeter

If the coils are properly connected and the scale properly calibrated, the meter will record power in watts.

Suppose that it is desired to measure the power supplied to a lamp *L*. The power coil is connected in series with the lamp. Practically all the current that goes through the lamp goes through the power coil, since the resistance of the voltage coil is relatively very high. The voltage coil is connected as a shunt across the terminals of the lamp, *BC*. The number of watts recorded by the meter is the power which is supplied to the lamp from the line *AC*.

A wattmeter will record power in watts from both d.c. and a.c. circuits.

453. The Watt-hour Meter. The watt-hour meter is in principle a combination of a wattmeter and an a.c. motor. The Thomson watt-hour meter, Fig. 436, is essentially a small induction motor whose field coils are connected like the current coil of a wattmeter and whose armature coils are connected like the voltage coil. The torque or turning effect acting on the armature is proportional to the power passing through the instrument, as in the case of the wattmeter.

The field coils of the instrument are connected in series with the power line, and the armature coil is con-

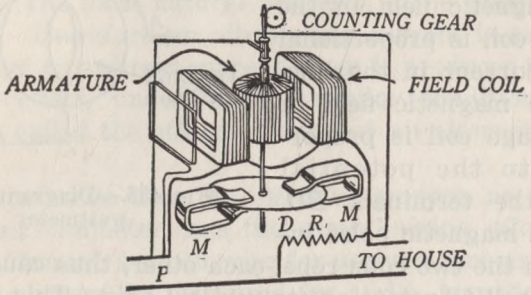


FIG. 436.—Watt-hour meter

nected in shunt across the power line. Thus we see that the connections both within the instrument and to the line are exactly the same as those described in connection with the wattmeter. The rotations of the armature are communicated by means of a worm gear to the recording dials, which register electrical energy in kilowatt-hours.

Attached to the lower armature shafts is a copper or aluminum disk which turns in a strong magnetic field, between the poles of two permanent magnets. The reaction of the eddy currents thus produced in the disc constitutes a drag on disc and armature which is proportional to the speed of the disc. Under these conditions the armature will turn at a speed proportional to the passing power, and the number of revolutions made by the armature in an hour, or in a day, will be proportional to the electrical energy which has passed into the connected circuit in that time. If it were not for the damping effect of the eddy currents in the disc, the machine, because of the inertia of its moving parts, would run too fast or it would fail to stop for some time after the current was cut off. We thus have here a very useful

application of the damping effect of eddy currents.

TRANSFORMERS

454. The Transformer. A *transformer* is a device for changing a current of high potential to one of low potential, or changing a current of low potential to one of high potential. The principle may be understood by considering Fig. 437. Suppose that an alternating current flows in the coil marked *P*, called the *primary*. The effect of this current in the primary is to magnetize the bar *B*. This will have the same effect as thrusting the magnet into the coil *S*, called the *secondary*. Let us suppose that the effect on the secondary is the same as if an N-pole had been thrust into it. When the current reverses in the primary, the polarity of the magnet is changed; that is to say, the N-pole becomes an S-pole. The effect of this change of polarity is the same as if the magnet were withdrawn, reversed with regard to its pole, and again thrust into the coil. Now it will be remembered that thrusting an N-pole into a coil gives rise to an indirect e.m.f., and withdrawing the magnet gives rise to a direct e.m.f. So, too, every time the current flows one way in the primary there is induced an e.m.f. in one direction, and every time the current reverses in the primary there is induced in the secondary an e.m.f. in the opposite direction. An alternating current in the primary produces an alternating e.m.f. in the secondary, and therefore an alternating current, provided the secondary circuit is closed.

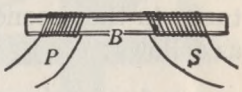


FIG. 437
Bar transformer

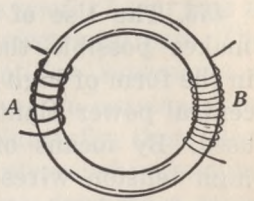


FIG. 438
Ring transformer

In Fig. 438 there is shown a form of transformer of the ring type. When A is the primary and B the secondary, the transformer is "step up"; when B is the primary and A the secondary, it is "step down." The relation between the e.m.f. of the primary and that of the secondary is determined by the relation of the number of turns of wire in the primary to the number of turns in the secondary. This relation may be expressed as follows:

$$\frac{\textit{e.m.f. of primary}}{\textit{e.m.f. of secondary}} = \frac{\textit{number turns of primary coil}}{\textit{number turns of secondary coil}}$$

Example. An alternating current having an e.m.f. of 2200 volts flows in the primary of a transformer having 200 turns in the primary coil and 10 in the secondary. Find the e.m.f. of the current in the secondary. *Solution*: Let x be the e.m.f. of the secondary. Then

$$\frac{2200}{x} = \frac{200}{10}, \text{ from which}$$
$$x = 110 \text{ volts}$$

This means that by the use of such a transformer as described in this example a high tension current having an e.m.f. of 2200 volts may be stepped down to a current having an e.m.f. of 110 volts.

455. The Use of the Transformer. The transformer makes possible the transmission of electrical energy in the form of high potential alternating currents from a central power plant to points where the energy is to be used. By means of the transformer a current on the high tension wires may always be stepped down to a current of low potential. Small commercial transformers, as shown in Fig. 439, are often attached to electric lighting poles adjacent to dwellings. The function of these transformers is of course to step the high potential

current down to one of low potential such as is safe to use in the lighting of houses, the running of domestic motors, and so forth.

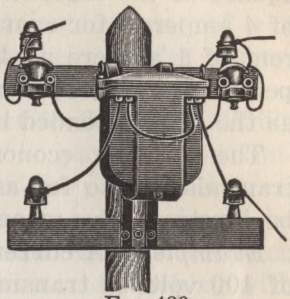


FIG. 439
Commercial transformer

The reason for using high tension currents on transmission lines is primarily one of economy in two respects. (a) When a very high e.m.f. is used, very little attention has to be paid to the resistance of the line. The effect of high resistance can always be overcome by using a still higher e.m.f. We may therefore use relatively fine wire for the high transmission purposes, line resistance not being an important factor. The use of relatively fine wires means a reduction in the expense of installing the system. (b) There is another reason also why a current of high voltage and low amperage is economical so far as transmission is concerned, and that is the fact that the heat loss for low current transmission is very much less than that for high current transmission. Let us take a long distance lighting current for example. The heat developed by the current may be divided into two parts, useful heat and waste heat. The heat developed in the lamps, giving rise to illumination, is useful heat; this is the energy for which consumers of electricity pay. The heat developed in the transmission line represents a dead loss; the power company receives no compensation for this wasted electrical energy. Naturally, then, electric light and power companies are interested in keeping the heat loss in the transmission line down to a minimum. Now we have learned (Art. 401) that the heat in calories developed in a conductor is $I^2Rt \times 0.24$; that is to say, for a conductor of given resistance and for a

given time the heat developed is proportional to the square of the current. The heat expended by a current of 4 amperes, for example, is to that expended by a current of 1 ampere as $4^2/1^2 = 16/1$; that is, the heat expended by a current of 4 amperes is 16 times as great as the heat expended by a current of 1 ampere.

The relative economy of high-tension low-current transmission, so far as line heat loss is concerned, may be illustrated by means of examples.

Example 1. A current of 10 amperes under a pressure of 100 volts is transmitted over a wire of 10 ohms resistance for 10 seconds. Find (a) the power of transmission and (b) the heat loss. *Solution*: Assuming that there is no inductance in the circuit, we have

(a) power $= EI = 100 \times 10 = 1000$ watts, and

(b) heat loss $= I^2Rt \times 0.24 = 10^2 \times 100 \times 0.24 = 2400$ calories

Example 2. Suppose now that the same power (Ex. 1) is transmitted by a current of 1 ampere under a pressure of 1000 volts over the same resistance for the same time. Find (a) the power expended and (b) the heat loss. *Solution*: Our equations now become

(a) power $= EI = 1000 \times 1 = 1000$ watts, and

(b) heat loss $= I^2Rt \times 0.24 = 1^2 \times 100 \times 0.24 = 24$ calories

For equal transmission of power in the cases cited above, the heat loss in the former case is 100 times as great as in the latter. We thus see why in high tension transmission we use currents of low amperage but of very high voltage. By means of a transformer a current of high voltage may be stepped down to one of low voltage and proportionally high amperage.

In the case of the transformer shown in Fig. 440, it should be noted that the wires carried by the upper cross piece of the pole are those of the high voltage circuit,

while those attached to the lower cross bar are those of the low voltage circuit. The low voltage wires are the ones which connect with the lighting systems of our dwelling houses, the current usually being under a pressure of 110 volts.

456. Avoiding Shocks from Electric Fixtures. While the current supplied to dwellings is generally not considered dangerous to life under ordinary circumstances, nevertheless a certain degree of care should always be exercised in handling electric fixtures. Electricians when standing upon a dry floor think nothing of touching with their fingers a 110 volt or even a 220 volt circuit to ascertain whether or not it is "alive." On the other hand one

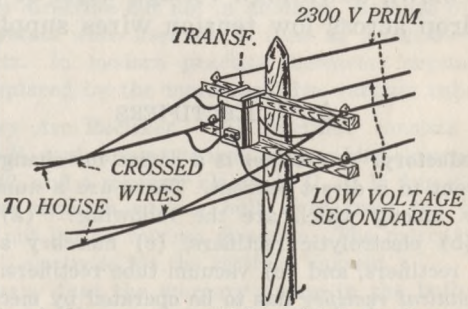


FIG. 440.—A dangerous condition: high tension primary wires crossed with low voltage circuit

should never touch the metal parts of an electric switch when the body is in contact with grounded lines, such as gas or water pipes. For example, one should never touch a switch while in a bath tub, as the shock received by the moist body may prove serious, or even fatal.

Also, dangerous and many times fatal shocks sometimes occur from the touching of electrical household fixtures due to the accidental crossing of high tension wires with those of the low tension circuit. How this may occur is shown in the accompanying illustration,

Fig. 440. Here is shown a high tension line (2300 volts) which, due to storm conditions, has sagged and come into contact with the low voltage circuit. Under this condition the fuses in the house will be blown as soon as the house circuit is closed; and further, the person who attempts to close a switch which is thus short-circuited with such a high tension wire, would be likely to receive a fatal charge. A number of cases are on record in which persons have lost their lives under conditions such as described.

A good point to keep in mind, therefore, is to exercise extreme caution in manipulating all lamp sockets or switches during or directly after a severe storm, which may have blown down high voltage wires so as to cause them to drop across low tension wires supplying house circuits.

A. C. Rectifiers

457. Introductory. A *rectifier* is a device for changing an alternating current to a direct current. There are a number of types of rectifiers among which are the following: (a) Mechanical rectifiers, (b) electrolytic rectifiers, (c) mercury arc rectifiers, (d) crystal rectifiers, and (e) vacuum tube rectifiers.

The *mechanical rectifier* has to be operated by mechanical power. The commutator on a d.c. generator is an example. The alternating current in the armature is changed by the commutator and brushes to a direct current. One of the largest rectifying units that we have is the rotary or synchronous converter (Art. 449) by means of which alternating currents from high tension lines are converted or rectified into a direct current by means of a d.c. generator which is attached to the shaft of the synchronous motor.

458. Electrolytic Rectifier. An *electrolytic* rectifier is a device for changing an alternating current to a direct current by the chemical action which takes place at certain electrode substances in an electrolytic cell. For example, it has been found that if one of the electrodes of an electrolytic cell consists of a plate of aluminum, this metal (aluminum) can act as a cathode but not as an anode. That is to say, if to an electrolytic cell having as an elec-

trolyte a solution of sodium phosphate or even a solution of common borax, and electrodes consisting of plates of lead (Pb) and aluminum (Al), Fig. 441, there be applied an alternating e.m.f., a current will flow through the cell from the lead electrode to the aluminum electrode, but not from the aluminum to the lead. It is clear that a large part of the electrical energy applied to a rectifier of this sort is lost, so far as useful work is concerned. One disadvantage arising from the use of an electrolytic rectifier is that the cell heats up very rapidly when in operation.

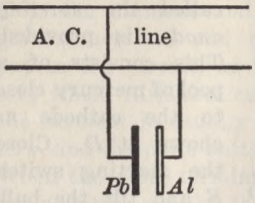


FIG. 441
Diagram of a lead-aluminum electrolytic rectifier

459. Crystal Rectifier. Certain crystals, especially those of carborundum, possess the property of transmitting a current in one direction but not in another. Because of this property such crystals were used for a time as detectors in wireless and radio sets. In modern practice, however, crystal detectors have been displaced by the more sensitive vacuum tube detectors.

460. Mercury Arc Rectifier. This rectifier consists of an exhausted bulb, *B*, containing two iron or graphite electrodes marked *G*, *G'*, Fig. 442, and a mercury electrode *C*. It is found that a current will pass through such a bulb only from the iron to the mercury, but not in the reverse direction. The mercury therefore becomes the + electrode for the rectified current.

It is necessary that the mercury vapor in the bulb be ionized in order to enable the current to pass from the iron electrode to the mercury electrode. A gas is said to be ionized when some electrons have been freed from the molecules. Under ordinary conditions gases are practically non-conductors of electricity. When a gas is ionized, however, and as a result contains a minute quantity of free electrons, it will conduct electricity freely.

In operating the mercury arc rectifier, the secondary terminals of an alternating current transformer are connected to the iron terminals of the rectifier. A wire connected to the center of the secondary of the transformer at *C* is attached to the negative terminal of the storage battery *SB*. The positive terminal of the battery is connected to the mercury electrode of the rectifier tube through a coil *R*, which by induction serves to sustain the arc between the alternations.

To start the rectifier, it is necessary first to ionize the mercury vapor in the bulb. To start the ionization an additional anode, called the *starting anode*, is provided. This consists of a pool of mercury close to the cathode as shown at *D*. Close the starting switch *S* and tilt the bulb first a little to the left and then back again. When the tube is tipped slightly the mercury *D* runs over and makes contact with *C*. Now, if the tube is tipped back, the mercury bridge connecting *C* and *D* is broken and the rupturing of the current causes a slight arc, *which is sufficient to produce the initial ionization of the mercury vapor.*

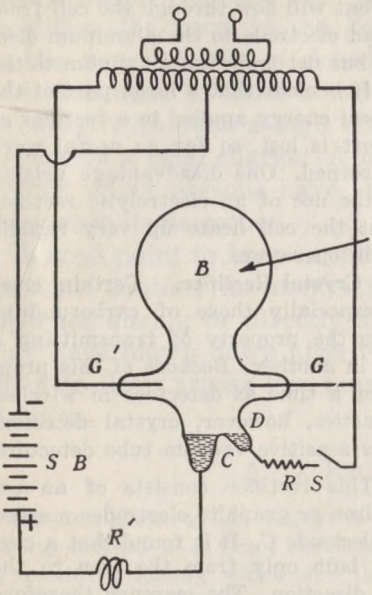


FIG. 442.—Mercury arc rectifier

The mercury arc rectifier is used quite extensively in garages to obtain direct currents for charging storage batteries. Tubes of large size will furnish direct currents as high as 20 to 30 amperes.

461. Vacuum Tube Rectifier. The *vacuum tube rectifier*, sometimes called a *vacuum valve*, much resembles an incandescent lamp. But, besides the filament, it contains a plate which sometimes takes the form of a metallic cylinder surrounding the filament. The discovery was made by Edison in 1883 that a current will flow in an evacuated bulb from the plate to the hot filament, but that it will not flow to any appreciable extent from the filament to the plate. In other words, the metal plate and the hot filament in the evacuated bulb constitute a rectifier, allowing a current to flow in one direction but not in the other. The current between the plate and the hot filament is carried by electrons which stream off from

the hot filament and impinge upon the plate. A fair understanding of the principle involved in Edison's vacuum valve discovery may be obtained from an inspection of Fig. 443, in which *F* is the filament and *P* the plate. A galvanometer *G* and two batteries are put in the circuit, as shown. The battery which furnishes current for heating the filament is usually called the *A battery*, and the plate battery the *B battery*. When the plate terminal is connected to the negative pole of the *B* battery, thus charging the plate *P* negatively, no current flows through the galvanometer, but when the plate is connected to the positive pole of *B* a current flows steadily through the galvanometer to the plate as long as the filament remains hot.

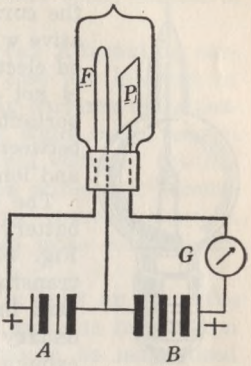


FIG. 443
Vacuum tube rectifier

A hot filament gives off electrons readily. When the plate has a — charge, the negatively charged electrons are repelled from it; when, on the other hand, the plate has a + charge, the electrons are attracted to it, thus giving rise to a current. We say, by convention, that the current flows through the galvanometer from the plate to the filament; what really occurs is that a stream of electrons flows from the filament to the plate.

Vacuum tubes are used extensively today as rectifiers, and as detectors and amplifiers in radio sets.

462. Thermionic Rectifiers. The *thermionic rectifier*, sold under the name of *tungar rectifier*, utilizes the principle of the mercury arc rectifier in addition to that of electron discharge from hot filaments. A tungar bulb is shown in Fig. 444. The bulb is filled with some inert gas, such as argon, at a pressure as low as from 3 to 8 centimeters of mercury, when cold. The anode *A* consists of a small cone of graphite or tungsten and the cathode *F* of a spiral of fine tungsten wire.

The anode is placed close to the tungsten filament, which can be heated to incandescence by passing a current through it. If a source of alternating voltage is connected to the anode and the hot cathode filament, a current will flow through the tube only when the filament is negative with respect to the anode. When the filament is negative, electrons which evaporate from its hot

surface are attracted to the anode and on their way across they ionize the argon. This ionized gas conducts the current quite freely. When the anode is negative with respect to the hot filament, the evaporated electrons are not attracted to it and so the gas is not ionized and does not conduct. It is important to note here that the current is conducted between the electrodes *A* and *F* by both electrons and ionized gas.

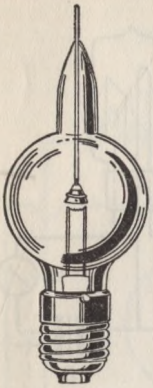


FIG. 444
Tungar bulb

The set-up of a tungar rectifier and a storage battery *SB* which is being charged is given in Fig. 445. The high tension side of a step-down transformer is connected to the alternating current lighting circuit, say of 110 volts. The secondary has a wire, tapped in, so that there will be enough voltage to heat to incandescence a short tungsten filament *F*. The voltage between *A* and *F* is not large enough to make the argon gas in the bulb conducting. But when the hot filament *F* emits electrons at a speed sufficient to begin to ionize the gas by collision, then the current flows through the tube from *A* to *F*. As the voltage from the transformer changes back and forth, first in one direction and then in the other, a current flows when *F* is negative and *A* positive; that is, half the time. The current, as in all rectifiers, is a pulsating one, but always in the same direction. A storage battery *SB* in series with the tube can therefore be charged.

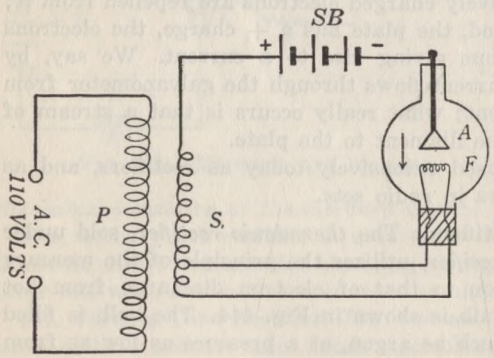


FIG. 445.—Tungar rectifier set

A medium sized tungar tube will carry a charging current of from 5 to 6 amperes.

SUMMARY

1. Direct currents are used (a) for electrolytic processes,

(b) for charging storage batteries, (c) for street railway service, and (d) for the operation of d. c. motors where speed control is an important factor.

2. Advantages of alternating currents: (a) Alternating currents may in general be produced at less cost per kilowatt of capacity than direct currents. (b) Alternating currents make possible the use of transformers by means of which high tension currents may be stepped down to any required voltage. (c) Because of the use of light weight transmission wires and because of reduced heat losses, power by alternating currents may be transmitted from power stations to consumers more economically than by direct currents.

3. In d.c. generators the magnetic field is built up from the small quantity of residual magnetism which remains in the iron cores; in a.c. generators the field cores have to be magnetized from a d.c. source, such as a battery or a small d.c. generator.

4. Magnetic fields may be stationary or rotary. The stationary part of a generator, whether field or armature, is called the stator; the rotating part is the rotor.

5. A cycle is a complete series of events during which an alternating electromotive force or current passes through zero going in a positive direction to zero again, going in the same direction. A cycle may be represented graphically by two consecutive loops of a curve, one up and the other down.

6. A cycle is divided into 360 electrical degrees.

7. Phase denotes the condition of change at any point in a cycle. Phase is measured in terms of electrical degrees.

8. Two alternating electromotive forces or currents are in step, that is in phase, when any point on one curve has the same number of electrical degrees as the corresponding point on the other.

9. A one phase generator delivers an alternating current which may be represented by a single curve.

10. A two phase generator delivers two alternating currents which differ in phase by 90°.

11. A three phase generator delivers three alternating currents which differ in phase by 120°.

12. Frequency is the number of cycles per second. For lighting purposes a frequency of 60 cycles per second is most commonly used; for power purposes, 25 cycles.

13. In d.c. circuits only three factors in general have to be

taken into account, namely, current I, electromotive force E, and resistance R; in a.c. circuits there may be as many as six factors, which are: current I, electromotive force E, resistance R, inductance L, capacitance C, and a power factor, the value of which depends on the lag or lead of the current with respect to the e.m.f.

14. The unit of inductance is the henry. A coil has an inductance of 1 henry when there is induced in it an e.m.f. of 1 volt by a current changing at the rate of 1 ampere per second.

15. The unit of capacitance is the farad. A condenser has a capacitance of 1 farad when a charge of 1 coulomb gives it a potential of 1 volt. A microfarad is a millionth of a farad.

16. The capacitance C of condensers C_1, C_2, C_3. . . in series is

$$\frac{1}{C} = \frac{1}{C_1} + \frac{1}{C_2} + \frac{1}{C_3}$$

17. The capacitance of C of condensers C_1, C_2, C_3. . . in parallel is

$$C = C_1 + C_2 + C_3 \ldots$$

18. Inductance in an a. c. circuit does two things to the current: (a) it chokes the current down, and (b) it causes the current to lag behind the e.m.f.

19. Capacitance in an a.c. circuit causes the current to lead the e.m.f.

20. Inductance may be varied (increased or decreased) by means of a variable induction coil; capacitance may be varied by means of a variable condenser.

21. Two alternating systems may be brought into phase (tuned) by varying L or C, or both.

22. The operation of an induction motor depends on the production of a rotating magnetic field. To produce a rotating field, it is in general necessary to have two or more currents which differ in phase.

23. A synchronous motor is of the nature of an a.c. generator running as a motor.

24. Ammeters and voltmeters containing permanent magnets such as are used in the measurement of d.c. units will not do for a.c. circuits. The essential parts of an a.c. ammeter or voltmeter is a coil which moves in a magnetic field in such a way that when the current reverses in the coil, the magnetic field reverses at the same time.

25. An a.c. ammeter records current values in effective amperes. An effective ampere is that current which will produce the

same heating effect in a conductor in a given time as will be produced by a direct current ampere. An a.c. voltmeter gives readings in effective volts.

26. In d.c. circuits, and in a.c. circuits in which L and C are negligible

$$power\ in\ watts = amperes \times volts$$

In a.c. circuits containing inductance or capacitance producing a lag or lead of the current

$$power\ in\ watts = amperes \times volts \times power\ factor$$

27. A wattmeter is an instrument for measuring electrical power in watts. Its essential parts are a power coil of low resistance and a voltage coil of high resistance. The torque (turning effect) produced by the magnetic field in the current coil is proportional to the current, and that of the voltage coil is proportional to the applied voltage. A wattmeter records power in watts for both d.c. and a.c. circuits.

28. A watt-hour meter is in principle a combination of a wattmeter and a small induction motor which drives a recording mechanism.

29. A transformer is a device for changing an alternating current from one voltage to another. When the voltage is stepped down, the current is stepped up, and vice versa. The power (EI) of the system remains unchanged.

30. The relation between e.m.f. and number of turns of wire is

$$\frac{e.m.f.\ of\ primary}{e.m.f.\ of\ secondary} = \frac{number\ of\ turns\ of\ primary\ coil}{number\ of\ turns\ of\ secondary\ coil}$$

31. A rectifier is a device for changing an alternating current to a direct current.

32. With the exception of mechanical rectifiers, such as rotary converters which are rectifiers on a large scale, all rectifiers depend in principle upon the fact that under certain conditions an alternating current will pass in one direction between two electrodes but not in the other. In rectification there is always a loss of electrical energy, which usually takes the form of heat.

REVIEW EXERCISES

A

1. Make a diagram to show the changes in e.m.f. when a single conductor makes one complete revolution in a magnetic field between an N-pole and an S-pole. Explain the character of the changes in each quarter revolution.

2. Explain briefly the difference between alternating and direct currents. Draw a d.c. curve; an a.c. curve.

3. How do you account for the fact that a d.c. generator will start without first having its field magnets excited by some outside magnetizing force?

4. Why do a.c. dynamos have to have their fields magnetized by an outside d.c. source of electricity?

5. Compare the respective advantages of direct and alternating currents.

6. By means of an a.c. curve explain what is meant by one cycle.

7. Explain the distinction between electrical and mechanical degrees.

8. Explain the meaning of the term phase. Draw curves to represent two alternating currents which differ in phase by 90°.

9. Why in general can't Ohm's law be applied to alternating currents?

10. What is the unit of inductance? What is the inductance of a coil in which a current changing at the rate of 2 amperes per second produces an induced e.m.f. of 1 volt?

11. What is the unit of capacitance? What is a microfarad?

12. What are the essential parts of an induction motor? By means of a diagram explain how a rotating field may be produced.

13. In an induction motor are there any slip rings or brushes? Does any of the line current directly enter the rotor?

14. Under what conditions will an a.c. generator act as a synchronous motor? Wherein does a synchronous motor differ from an induction motor?

15. What is a rotary converter, and for what is it used?

16. So far as essential parts are concerned, wherein does an a.c. ammeter differ from a d.c. ammeter?

17. In general, how may one tell by looking at the scale of an ammeter whether it is a.c. or d.c.?

18. On d.c. ammeters and voltmeters one binding post is marked plus (+) and the other minus (—). Is this necessary on a.c. instruments? Why?

19. What are the essential parts of a wattmeter? Will a wattmeter measure power on a d.c. circuit?

20. What are the essential parts of a watt-hour meter? What does such an instrument record?

21. Is a watt a unit of power? Is a watt-hour a unit of power? What is a kilowatt-hour?

22. Make a sketch of a simple transformer, lettering one coil *A* and the other *B*. Which coil is the primary and which the secondary when the transformer is used to step the e.m.f. (a) up? (b) down?

23. A 110-volt lighting system is supplied through a step-down transformer from a high tension circuit, the voltage of which is 1100. What must be the ratio of the number of turns of the primary coil to the number of turns of the secondary?

24. It is desired to step a 220 volt current down to one of 10 volts. If there are 440 turns in the primary, how many turns must there be in the secondary?

25. What is an electrical rectifier? Name some of the more important types of a.c. rectifiers.

26. (a) In a lead-aluminum electrolytic rectifier, in which direction does the current flow through the cell, toward the lead or the aluminum? (b) What becomes of that part of the electrical energy which does not pass through the cell as a current?

B

27. What is a magneto? Can you think of any uses to which this type of dynamo is put?

28. Explain the meaning of the terms "stator," "rotor."

29. What are some of the advantages of using rotating magnetic fields in a.c. generators?

30. Make a sketch to illustrate a two-phase generator and its circuits.

31. How many wires are used in a three-phase circuit?

32. Given two condensers having capacities of 0.4 and 0.6 microfarad respectively. Find the capacitance when these condensers are connected (a) in series, (b) in parallel.

33. What are the effects of inductance in an a.c. circuit?

34. A given coil is connected to a d.c. line having a potential of 110 volts. A d.c. ammeter connected in series with the coil indicates a current of 5 amperes. The coil is now connected to an a.c. line of 110 volts. An a.c. ammeter shows that a current of 0.5 ampere is flowing through the coil. How do you account for this difference in current strength in the two cases?

35. It is found that due to inductance in a given a.c. circuit,

the current lags behind the impressed e.m.f. by 30 electrical degrees. Draw curves to represent this condition of lag.

36. An electric lamp is put in a d.c. circuit in which a condenser is connected in series. Will the lamp light up when the circuit key is closed? Explain.

37. In a given a.c. circuit there is a coil of fixed inductance and a variable condenser. The inductance causes the current to lag behind the e.m.f. How in this case can the current be brought into phase with the e.m.f.?

38. The electric power delivered to a line of 2 ohms resistance in each of the following cases is 50 kilowatts. Case (a): The voltage at the power house end of the line $E = 500$ volts; current $I = 100$ amperes; $R = 2$ ohms. Case (b): $E = 10{,}000$ volts; $I = 5$ amperes; $R = 2$ ohms. (a) Find the I^2R heat loss in watts in each case. (b) What are the respective advantages of high voltage and low amperage transmission?

C

39. Describe the flow of electrons in the case of an alternating current.

40. An electric current under a pressure of 110 volts is furnished to a building containing lamps connected in parallel. On a certain evening 100 lamps, each taking 0.5 ampere, were lighted for 4 hours. (a) What power in watts was furnished to the lamps? What was the total energy consumed (b) in watt-hours? (c) in kilowatt-hours?

41. (a) Make a diagram of a mercury arc rectifier and explain its operation. (b) When is a gas said to be ionized? (c) What about the conductivity of a gas that is ionized? (d) Why is it necessary to tilt the rectifier in order to start it working?

42. Make a diagram of a vacuum tube rectifier and explain its operation. Explain why the current flows from the hot filament to the plate when the latter is positively charged and not when it is negatively charged.

43. (a) What is the function of the A battery in a vacuum tube rectifier? (b) the B battery?

44. (a) Explain by means of a diagram how a storage battery may be charged by means of a tungar rectifier. (b) What conducts the current in a thermionic rectifier?

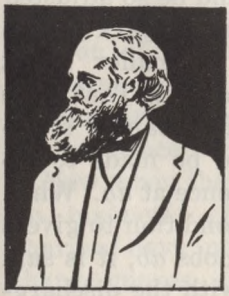
NOTE: For additional problems, see Appendix.

CHAPTER XV

ELECTROMAGNETIC WAVES

Properties of Electric Waves

463. Introductory. It is assumed that all space is filled with a highly elastic and weightless medium called the *ether*, which not only occupies the space between the heavenly bodies but also pervades the atomic and molecular structure of matter. In the neighborhood of a charged body the ether acts as if it were under a state of strain. It is also assumed that the ether in a magnetic field is in a strained condition. It follows that upon the sudden discharge of a charged body or the sudden demagnetization of a magnet, a disturbance of the ether takes place. Such a disturbance of the ether is called an *electromagnetic wave*, since both electrostatic and magnetic effects are present.



Maxwell

In 1864 Maxwell, a noted English physicist and mathematician, advanced a theory that by means of electrical disturbances it ought to be possible to create electromagnetic waves in the ether similar to light waves. Later on this theory was confirmed experimentally by Hertz.

464. Hertzian Waves. In 1887 Hertz, a German physicist, discovered that electrical oscillations, such as occur on the discharge of a Leyden jar, give rise to electrical

waves in the ether. Thus when the oscillatory discharge occurs between the knobs of an electric machine or between the terminals of an induction coil, electric waves are set up in the ether and travel out in all directions from the source. These waves, which came to be known as Hertzian waves, are capable of being reflected, refracted, and polarized the same as light waves. In fact the electric waves seem to possess all the properties of light waves, differing only in the fact that they are very much longer than those of light.



Hertz

In order to detect these electrical waves, Hertz used a device called a spark-gap detector, Fig. 446. This consists of a rectangular wire C set up near the poles of an electric machine or induction coil. It was found that the size of the loop C could be so adjusted by means of the sliding wire S that when a spark passed between AB, a smaller spark would appear at ab. That is to say, electric disturbances started at AB could be made to record their presence at ab. When the loop is in condition to give a spark at the knobs ab, it is said to be *in tune* with the discharging system AB.

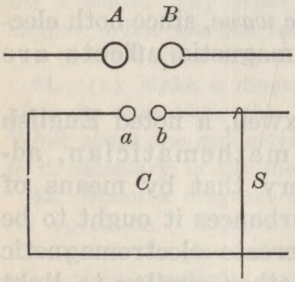


FIG. 446
The Hertz spark-gap detector

This discovery by Hertz contained the fundamental principle of signaling through space by means of electric waves.

465. The Nature of Electromagnetic Waves. Let AB, Fig. 447, be the discharge terminals of a condenser

connected with an electric machine or induction coil. When the potential rises sufficiently high, a spark passes between the knobs AB. Just before the discharge takes place, the highly charged rods are surrounded by an electrostatic field, which extends in all directions about the oscillator. For convenience the lines of the figure are represented on one side of the rods only. When the discharge occurs, there will be a sort of inertia effect. This rush of discharge will go too far, resulting in a reversal of the charges in the rods. This is followed by another discharge in the opposite direction, and so on, for a series of back and forth oscillations.



FIG. 447

When the discharge of the oscillator occurs, the field near the rods disappears momentarily, reappearing again, but reversed in direction. The resurge of the first oscillation recharges the rods in the opposite sense. The portion of the field at C passes outward into the surrounding space. One can imagine that at the moment of discharge the ends of the lines of force are snapped off into space.

Now when the charges on the oscillator are stationary, no magnetic effects are produced, since stationary charges have no magnetic fields. But as soon as the charge moves, as in a series of oscillatory discharges, magnetic fields are produced, because a moving charge is an electric current and is surrounded by a magnetic field. Thus it appears that electromagnetic waves consist of a combination of moving oscillatory magnetic and electrostatic fields.

466. Velocity, Wave Length, and Frequency. Electromagnetic waves travel through space with the speed

of light, namely, 186,000 miles or 300,000,000 meters per second. In Fig. 448 there is pictured a series of waves moving in the direction AB. That portion of the curve marked $abcde$ represents one wave or cycle. The *amplitude* of vibration for any given wave is its greatest vertical distance above or below the axis, as bb' and dd'. *Wave length* is the distance from any point on one wave to the corresponding point on the next wave. The *period* T is the time of one complete vibration. *Frequency* is the number of vibrations per second.

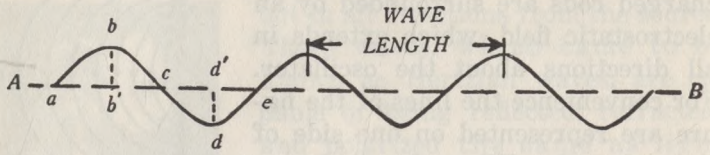


FIG. 448
Wave length, amplitude, and frequency

It is very important to have a clear idea of the relation between period, frequency, wave length, and velocity. As an illustration of the relation between period and frequency let us consider again the vibration of a simple pendulum (Art. 174), the periodic motions of which correspond to those of a wave. Suppose that the period of a given pendulum is 1/4 second. Then it is clear that in one second the pendulum will make 4 vibrations. That is, its *period* = 1/4 second, and its *frequency* = 4. Frequency is equal to the reciprocal of the period, that is,

$$\textit{frequency} = \frac{1}{\textit{period}} = \frac{1}{T}$$

Also, since the total distance traveled by a wave train in one second is equal to the distance traveled by one wave multiplied by the number of waves, we may write

$$\textit{velocity} = \textit{wave length} \times \textit{frequency}$$

But since the velocity of electromagnetic waves is 300,000,000 meters per second,

$$300{,}000{,}000 = \textit{wave length} \times \textit{frequency}$$

from which we have for electromagnetic waves

$$\textit{wave length} = \frac{300{,}000{,}000}{\textit{frequency}}$$

$$\textit{frequency} = \frac{300{,}000{,}000}{\textit{wave length}}$$

For example, if the wave length of a given radio wave is reported as 300 meters, then its

$$\textit{frequency} = \frac{300{,}000{,}000}{300} = 1{,}000{,}000$$

The U. S. Bureau of Standards at Washington has recommended to broadcasting corporations that wave frequency be expressed in "kilocycles," the corresponding wave length in meters being given in parentheses. For example, the term "1000 (300)" means a radio wave train having a frequency of 1000 kilocycles and a wave length of 300 meters. A *kilocycle* is 1000 cycles, that is,

$$\textit{kilocycles} = \frac{\textit{frequency}}{1000}$$

467. Wireless Telegraph. The wireless telegraph to a great extent has been superceded by the wireless telephone of our modern radio service. The subject is treated here, however, first, because of its historic interest, and second, because it serves as an appropriate introduction to the subject of Radio which is to follow.

The researches by Hertz on electrical oscillations opened up the way for the transmission of messages by wireless. The idea of the transmission of messages by means of electrical waves was developed and perfected by Marconi, a brilliant young Italian physicist, who invented the wireless system which became the forerunner of our modern "radio."

When a spark passes between the knobs of an induction coil, a violent rush of electricity occurs, which may be accompanied by electric oscillations in any conductors connected with them. This electric surging gives rise to an electric disturbance which spreads out through space. Now when these electric waves reach a distant conductor, they cause a somewhat corresponding disturbance in the conductor.



Marconi

In 1890 Branly and Lodge prepared a delicate detector for these electric disturbances, now known as Hertz waves. This detector is called a *coherer,* Fig. 449. It consists of a glass tube loosely packed with metal filings. When the coherer is placed in an ordinary battery circuit, it offers such a high resistance that scarcely any current will pass through it. It was discovered, however, that if an electric spark be produced in a near-by coil, the filings in the tube at once conduct readily and continue to act as a good conductor until the tube is disturbed by tapping. Marconi seized upon this discovery and after much experimentation devised an apparatus which would respond to the Hertz waves over a distance of hundreds of miles.

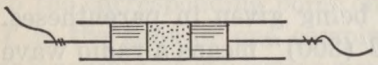


FIG. 449.—The coherer detector

In the development of modern radio systems a variety of detectors were successfully used. These may be classified in order as (a) coherers, (b) magnetic detectors, (c) thermal detectors, (d) electrolytic detectors, (e) crystal detectors, and (f) vacuum tube detectors.

Practically all of the earlier types of detectors have been replaced by the vacuum (electron) tube detectors.

A wireless telegraph system consists of a sending set and a receiving set. The essential parts of the sending set are an induction coil which is operated by means of a key, and a sending aerial or antenna. Let us suppose that 1000 sparks per second are produced by the sending oscillator, each spark sending out waves having a frequency of 1,000,000 vibrations per second. Then when the key is closed, a series of 1000 wave trains per second move out into space with the velocity of light.

The receiving set includes a receiving aerial or antenna *A*, Fig. 450, a detector *R*, and a telephone *T*. The detector pictured in Fig. 450 is of the crystal rectifier type. Like the vacuum tube rectifier, and others, it has the property of transmitting a current in one direction only. It is important to note the use of the detector. Suppose the oscillations in the receiving circuit have a frequency of 1,000,000 per second. If these currents could pass through the telephone, and if it were possible for the diaphragm of the instrument to respond to them, no effect would be produced upon the listening ear, since the ear is incapable of hearing sounds of a frequency above 20,000 or 30,000 per second. Hence, without the aid of the crystal detector, or its equivalent, the telephone would give no evidence of the presence of the oscillating discharges. If these oscillating currents which are induced in the receiving antenna system are rectified, however, each wave train will produce an effect like that of a *single pulse* in one direction. To these rectified wave trains the telephone receiver responds, and the listening ear hears a tone having a

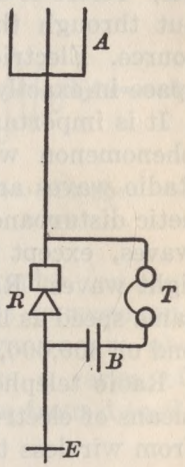


FIG. 450
Diagram of a crystal receiving set

pitch which corresponds to 1000 vibrations per second. As the key at the sending station is closed and opened, the series of wave trains is broken up into short and long groups which produce in the receiving telephone a series of short and long notes corresponding to the dots and dashes of ordinary telegraphy.

RADIO

468. Introductory. One of the most interesting and important of the modern developments of electric science is that which is now universally known as "Radio." The term radio, which is derived from the word "radiate," refers to an electrical disturbance which is radiated out through the ether in all directions from a given source. Electrical radio waves are transmitted through space in exactly the same way as are light waves.

It is important to note that radio is an electrical wave phenomenon which is transmitted through the ether. Radio waves are not sound waves; they are electromagnetic disturbances, having all the characteristics of light waves, except that they are enormously longer than light waves. Radio waves travel through space with the same speed as light waves, namely, 186,000 miles per second or 300,000,000 meters per second.

Radio telephony is the art of transmitting speech by means of electromagnetic waves. Radio telephony differs from wireless telegraphy in certain very important particulars. So far as reception is concerned, wireless telegraphy is simply a series of short and long notes (dots and dashes) of a given pitch. In radio it is necessary to vary the amplitude of the waves so as to produce the various tones of the human voice. The essential conditions for radio telephony may be stated briefly as follows: (a) the oscillator at the sending station must be so designed as to send out an *unbroken succession of*

waves of uniform intensity; and (b) in place of a key, such as was used in the wireless sending station for starting and stopping the oscillations, there must be substituted a telephone transmitter connected to the sending antenna in such a manner that the talking currents in the transmitter circuit will cause the oscillations to vary in intensity, thus producing *modulated tones* in the receiving apparatus.

469. Kinds of Radio Waves. Radio waves are of three kinds; damped, continuous, and modulated. When a spark occurs between the terminals of a Hertz oscillator, for example, there is sent out a wave motion in the ether, Fig. 451, the vibrations of which are said to be *damped*; that is, the vibrations die out quickly, each one being less in amplitude than the one preceding it. This damping effect is due mainly to resistance encountered by the oscillating current in the spark gap. Damped oscillations are suitable for wireless telegraphy, *but cannot be used in radio telephony.* For this purpose undamped oscillations giving continuous waves are obviously necessary. In Fig. 452 we have a series of damped wave trains moving in the direction from *A* to *B*.

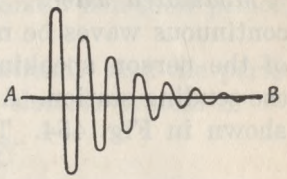


FIG. 451.—Damped wave

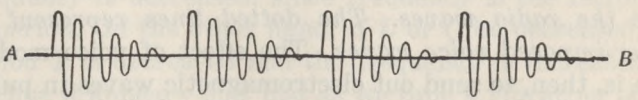


FIG. 452.—Train of damped waves

Continuous waves. The first necessary condition for the reproduction of the tones of the human voice is that the oscillator at the sending station shall send out a con-

tinuous series of waves of uniform intensity, Fig. 453. These are sometimes called *carrier waves.* The frequency of such waves, however, is so great as to make it impossible for the telephone receiver to respond to them, and

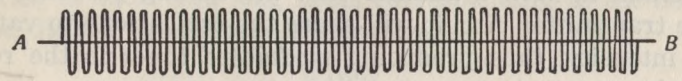


FIG. 453.—Undamped wave train

even if the telephone could respond, no sound would be heard because the audibility range of the human ear is not great enough.

Modulated waves. The second condition is that these continuous waves be modulated in intensity by the voice of the person speaking into the transmitting device at the sending station. A series of voice modulated waves is shown in Fig. 454. The frequency of the human voice

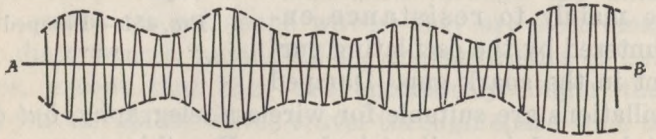


FIG. 454.—Modulated wave train

is very low as compared with the radio frequency. Many radio waves are therefore included in, and affected by each wave of the voice current. *The light lines represent the radio waves. The dotted lines represent the super-imposed voice waves.* The effect of voice modulation is, then, to send out electromagnetic waves in pulses, the frequencies of the pulses being low enough to be heard by the person at the receiving end.

The number of vibrations per second of the light lines, Fig. 454, is the *radio frequency*; the number of vibrations per second of the dotted lines, the *audio frequency.* The

radio frequency may be as high as 1,000,000; the audio frequency in general lies well below 20,000, usually in the neighborhood of 1000.

470. Tuning. Let us consider now how a receiving set is tuned to any given frequency. Two vibrating strings, as for example those of a violin or guitar, are in tune when their frequencies are the same. If the frequency of one string is fixed, the second string may be tuned to the first by changing the length or the tension of the second. In an analogous manner, two vibrating electric systems may be tuned by varying the inductance L or the capacitance C of one of the systems. The frequencies of two vibrating systems are the same when their periods are the same.

Now it may be shown mathematically that the period of a simple harmonic oscillating electric system is

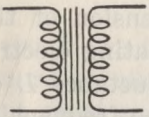
$$T = 2\pi \sqrt{L\,C}$$

where T is the period in seconds, L the inductance in henrys, and C the capacitance in farads.

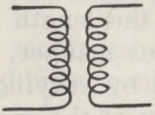
The significance of this equation lies in the fact that it tells us that the period of vibration of an oscillating electric system, such as described, is dependent upon the factors of *inductance* and *capacitance*, both of which may readily be varied. If we increase L or C, or both, the period T of the system is increased, and consequently the frequency is decreased, since frequency is the reciprocal of period. On the other hand, if L or C be decreased, the period T is decreased and the frequency is increased.

Thus it appears that just as we tune a vibrating string by changing its length or tension, so, too, we may tune a receiving set by changing the inductance or capacitance of the antenna section of the system. In practice, this is most frequently done by changing the capacitance. For this purpose, receiving sets contain condens-

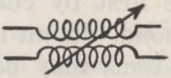
ers, the capacitance of which may be varied by rotating the plates forming one pole of the condenser. These variable condensers are usually connected in parallel with other parts of the set, taking advantage of the fact that increasing the capacitance increases the period of vibration and the wave length. Condensers connected in series would have the opposite effect; they are so con-



AUDIO FREQUENCY TRANSFORMER



RADIO FREQUENCY TRANSFORMER



VARIABLE COUPLING OF COILS

FIG. 455

nected in sets for receiving waves of short length or period.

The conventional symbols for inductance coils and condensers are given in Figs. 455 and 456.

471. Essential Parts of the Receiving Set. Apparatus for radio service includes primarily a broadcasting set (to be discussed later) and a receiving set. The essential parts of every receiving set, as shown in Fig. 457, include the following: (a) Antenna, a wire or group of wires without covering, which must be elevated from the ground and which may be located either out of doors on poles or may be suspended from insulators inside the building. Sometimes a radio loop is substituted for the outdoor antenna. (b) Lead-in wire, which should always be soldered to the antenna. (c) Ground connection, which is usually grounded by attaching to a steam or water pipe. (d) Tuning set for adjusting the receiving



FIXED CONDENSER



VARIABLE CONDENSER

FIG. 456

set to different wave lengths, so that radio waves of any desired frequency may be intercepted. (e) Detector for

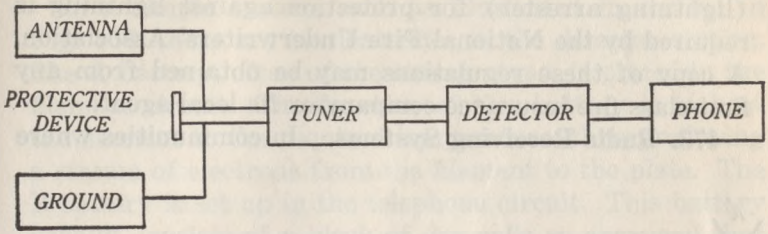


FIG. 457.—Essential parts of a radio receiving set

changing the frequency of the radio waves at the receiving station from high frequency to a lower frequency which may be heard.

472. Antenna Installation. Antennas may, in general, be classified as (a) receiving, and (b) transmitting. Receiving antennas are of two general types, namely, outdoor and indoor. For an antenna of the elevated outdoor type, a single bare wire is usually most satisfactory. A very excellent form is that which consists of a single wire about 75 to 100 feet long and elevated from 30 to 50 feet above the ground. In general, the greater the height, the better. If the antenna must be shorter than 75 feet, two parallel wires should be used, spaced about 3 feet apart, and joined together at one end of the loop.

All wire connections should be soldered, except where such connections are made by means of nuts and screws which may be turned up so tightly as to bite into the metal and thus make perfect electrical contact.

Radio transmitting stations have to be equipped with a more complicated antenna system than is required for receiving stations. Transmitting antennas (sometimes called aerials) usually consist of a system of long paral-

lel wires which have to be insulated very carefully because of the high voltages employed in sending.

In the installation of antennas an approved device (lightning arrester) for protection against lightning is required by the National Fire Underwriters' Association. A copy of these regulations may be obtained from any first-class fire insurance company or its local agent.

473. Radio Receiving Systems. In communities where

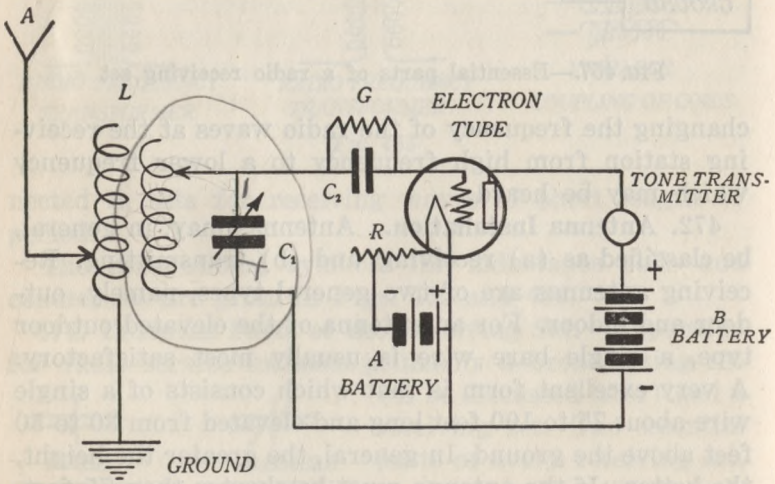


FIG. 458.—Detail of a receiving set

the usual alternating current house-lighting system is in use, a.c. radio sets have in a large measure displaced the older d.c. battery type. We shall, however, begin our study of this subject with a discussion of the battery receiving set, not only because it furnishes a natural historic approach, but also because it makes an explanation of the fundamental principles relatively simple.

In Fig. 458 there is shown in outline the principal

parts of a simple set-up of a receiving system. Antenna A is connected to the inductance system L, and thence to the ground.

In a simple system, such as illustrated in outline in Fig. 458, there are two batteries in the detector section. The A battery, often of the storage type, is placed in the filament circuit. The function of this battery is to heat the filament to the required temperature, thus producing a stream of electrons from the filament to the plate. The B battery is set up in the telephone circuit. This battery usually consists of a block of dry cells so arranged that the plate voltage may be varied by changing the battery connections. Note that the B battery is connected up with the telephone circuit so that its positive pole is in contact with the plate. The negatively charged electrons flow from the hot filament to the positively charged plate.

The variable inductance coils L and condenser C_1 are for the purpose of tuning. The use of the grid leak G and the condenser C_2 will be discussed later. We shall now take up for consideration the most essential element in the radio set, namely, the electron tube.

474. Electron Tube Detector. The principle of the electron or vacuum tube detector was discovered by Edison (Art. 461). Modern wireless telephony, or radio as it is commonly called today, was made possible, however, by De Forest's invention of the "audion" which is a modification of the Edison vacuum tube. The *audion* is an exceedingly sensitive vacuum tube containing, in addition to the filament and plate, a "grid," which consists of a zig-zag wire sealed within the bulb and placed *between* the filament and the plate, Fig. 459.

De Forest found that by means of his audion the volume of telephonic voice currents could be amplified many times, so that loud telephonic speech could be obtained

by voice currents which were originally too faint to be heard distinctly. Further, it was found that when two or more of these electron tubes are connected in a telephone circuit so that the amplified current of the first tube goes to the second, and from the second to the third, the telephone transmitter becomes microphonic; that is, exceedingly faint sounds may be amplified several hundred to a thousand times in magnitude.

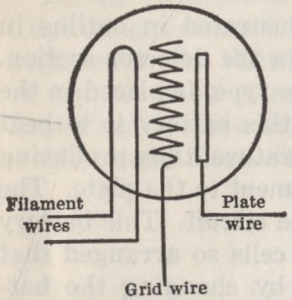


FIG. 459.—Diagram of an electron tube

The function of the electron tube, Fig. 459, is to act as a sort of electric valve which is actuated by the weak currents in the antenna system and which opens and closes the battery circuit of the telephone system. The electron tube acts in somewhat the same manner as does the relay of a telegraph system.

475. The Function of the Grid. The magnitude of the current in the plate circuit of the vacuum tube depends upon several factors, which we shall now consider in order. First, for a given temperature of the filament and a given voltage applied to the plate, the current in the B circuit will be greater the closer the plate is to the filament.

Second, for a given bulb the plate current varies with the voltage of the B battery, and with the temperature of the filament.

There is a third important factor upon which the current in the plate circuit depends and this is the potential of the space between the plate and the filament. As electrons leave the hot filament and move toward the plate, they give what is called a "space charge" to the vacuous space in the vicinity of the filament. This space charge

is negative and it therefore tends to limit the flow of electrons from the filament to the plate. Here is where the use of the grid comes in. If a grid be introduced into the bulb and placed close to the filament, this space charge may be controlled by varying the potential of the grid. If the grid is positive, its charge will tend to neutralize the space charge which is negative and consequently there will be an increased flow of electrons from the filament to the plate. If, on the other hand, the grid is negative, it will repel the negatively charged electrons and will therefore tend to decrease their flow. It follows, then, that if the voltage applied to the plate is kept constant and the temperature of the filament is also kept constant, the current in the plate circuit depends upon the potential of the grid. The flow of electrons may be entirely checked by the grid when it is negatively charged, and on the other hand the flow may reach a maximum when it is positively charged. Between these two extremes we may have varying conditions of grid potential producing varying currents in the plate circuit.

The potential of the grid is controlled either by a special battery known as the *C* battery, or by means of the *A* battery; this latter is the most common means of giving the grid its initial potential.

476. Grid Leak and Grid Condenser. The action of the grid leak and grid condenser, *G* of Fig. 458, is as follows: The incoming signals make the terminals of the condenser *C* alternatively positive and negative. This produces an alternating voltage across the condenser in *G*, thus causing the grid to become alternately positive and negative. When the grid is positive, electrons are attracted to it; when negative, no electrons are attracted. Consequently, due to the alternating voltage applied, electrons are attracted to the grid only during the positive half of the wave. The electrons accumulate on the grid,

giving it a preponderant negative charge, and thus tending to decrease the plate circuit.

If the vacuum of the bulb is low, this negative charge will leak off the grid slowly between sparks. If the vacuum of the bulb is high, however, the negative charge will not leak off through the bulb, but will maintain the grid at a negative potential. To avoid this, a high resistance is placed across the condenser in *G*. This resistance is not low enough to permit the condenser to discharge as fast as it is charged; but on the other hand, it is low enough to permit the condenser to discharge between sparks.

The function of the grid leak across the grid condenser, then, is to prevent an undue accumulation of negative charges in bulbs having a high vacuum.

477. Radio Broadcasting. The apparatus employed for transmitting radio messages differs from that used for receiving purposes, mainly in the arrangement by which the continuous current produced by the generator circuit is modulated to follow the variations of the voice. In transmitting sets, the various parts of the circuit must be designed to carry considerably larger currents than those employed in receiving sets. As in receiving sets, the filament of the electron tube used in transmission is heated by a current from a storage battery. The B battery, however, is replaced by a dynamo which is capable of supplying more power than can be economically obtained from batteries.

478. Amplification and Loud Speakers. Since a small varying grid current in an electron tube releases very much larger currents in the plate circuit, it follows that every electron tube is of the nature of a step-up current transformer. Now this magnified output from the first tube may be passed into a second tube, thus producing a still further amplification. A series of electron

tubes connected up in this way constitutes, with their accessory connections, what is known as an *amplifier*. Receiving sets are usually equipped with a device called a *loud speaker* which is placed in the telephone section of the circuit and which makes use of a horn, megaphone, or phonograph sound box to concentrate and direct the sound. Loud speakers, in general, may be divided into two classes: (a) The sound distributor type which uses an ordinary telephone receiver as a part of the equipment, and (b) the sound producer type which is used where extremely loud sounds are required.

479. A.C. Radio Sets. Alternating current radio sets are designed to operate directly from the usual 110-volt a.c. lighting socket, thus eliminating all battery connections. Developments in nearly every field of human endeavor have been preceded by public demand, and the a.c. radio set is no exception to the rule. Even in the early days of radio, the set owners grumbled about battery troubles. As a result, B eliminators were introduced and perfected, but the storage A battery still remained to cause bother and annoyance. "Trickle chargers" helped the situation somewhat, but the advent of the new a.c. tubes has now made it possible to operate the radio receiver direct from the lamp socket without the help, or hindrance, of any batteries at all. With full a.c. electric operation it is possible to eliminate entirely the necessity of testing, charging, and filling the storage A battery, and of replacing the B batteries at regular intervals. In fact, one gets rid of every attention and replacement required by the battery-operated set except the replacement of the vacuum tubes. On the other hand, a.c. tubes are more expensive than those used in the battery-operated set and in general do not last quite so long.

Essentially, conversion to full electric operation means, first, substituting a.c. type tubes for the storage battery

type tubes, and a step down filament-heating transformer for the storage battery. The alternating current for the filament supply is obtained from the a.c. lighting line through the use of a small transformer to step the voltage down to the pressure required.

In the second place, it is necessary to provide means for changing the alternating current to a direct current for the B battery. This is done by the use of a vacuum tube rectifier, which, with its accessories, is called a "B eliminator." The quality of any B eliminator is measured by how nearly it comes to supplying a perfectly smooth direct current and by its ability to maintain a constant voltage, regardless of the amount of current drawn from it or the fluctuations in the voltage of the electric light current that is used to operate it. If the pulsations are not removed from the current by the time it goes from the binding posts of the eliminator, a certain amount of hum will be produced in the loud speaker. The hum, as indicated by the alternating current voltage still remaining in the output of the eliminator, has been reduced, however, to an extremely low figure.

480. Television. *Television* or "radio sight" is the process of transmitting pictures by radio. As in voice transmission, there is a sending set and a receiving set. The essential parts of the sending set, Fig. 460, consist of a powerful source of light for the illuminating the object to be "televised," a perforated rotating disc, a photo-electric cell, and a radio transmitter.

The perforated disc, called the *scanning disc,* contains a series of holes, each a little nearer the center than the one preceding it. The scanning disc is firmly mounted on the shaft of a motor and is rotated at a fixed speed. The object of the disc is to chop the light beam up into patches of light which skim over the face of the sub-

ject, until the entire surface has been "scanned." The sensation of vision in the human eye persists for the sixteenth of a second or thereabouts. In moving pictures

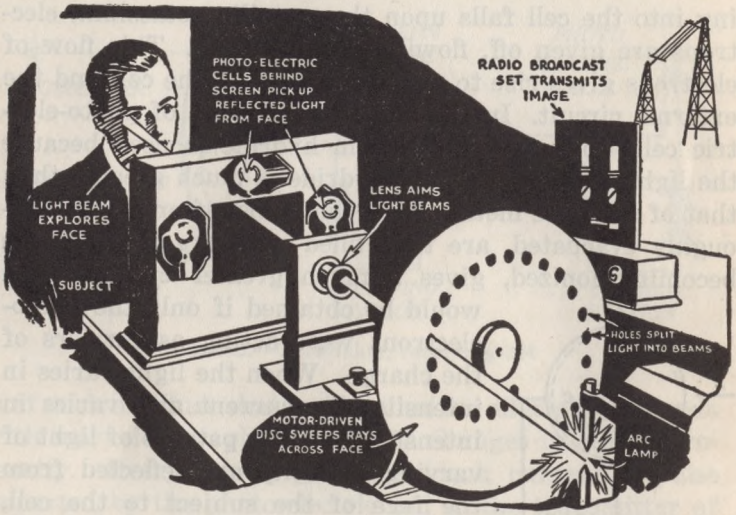


FIG. 460.—Television, sending set

the photographs are projected on the screen at the rate of sixteen per second, and the ability of the eye to retain vision for a fraction of a second makes the pictures blend together. Thus we get the effect of continuous motion. So, too, in television, if we are apparently to see motion, we must produce a series of individual pictures or images, and transmit them to the eye with about the same rapidity that motion pictures are thrown on the screen.

The *photo-electric cell,* Fig. 461, consists of a bulb in the center of which there is a loop of wire *A,* the anode. Over most of the interior of the bulb there is a deposit

of metallic sodium or potassium which constitutes the cathode *C*. On one side of the cell there is a clear space through which the light enters. The anode *A* is connected to the positive pole of the battery and the cathode *C* is connected to the negative pole. When the light passing into the cell falls upon the metallic potassium, electrons are given off, flowing from *C* to *A*. This flow of electrons gives rise to a current through the cell and the external circuit. In the more recent types of photo-electric cells sodium or potassium hydride is used because the light sensitivity of the hydride is much greater than that of the pure metal. Also, the bulbs, after being thoroughly evacuated, are then filled with argon which, on becoming ionized, gives a much greater current than would be obtained if only the photo-electrons were acting as carriers of the charge. When the light varies in intensity, the current also varies in intensity. Shifting patches of light of varying intensity are reflected from the face of the subject to the cell, which responds by sending corresponding varying current impulses to the transmitter. These impulses are amplified and radioed into space from the sending aerial. In high power transmission sets several photo-electric cells are used.

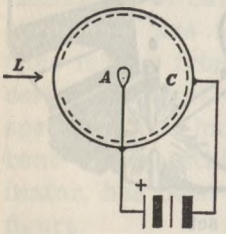


FIG. 461
Photo-electric cell

The essential parts of the receiving set, Fig. 462, are a perforated disc similar in all respects to the one used in the sending station, a neon-lamp, and an amplifier. The receiving disc contains the same number of holes spaced in exactly the same way as the sending disc. Moreover the receiving disc must rotate at the same speed as the sending disc. The speed of the receiving disc is regulated by a push button in the hand of the receiving

operator. This speed regulation corresponds in a way to tuning in ordinary radio reception.

The *neon-tube* is an incandescent electric lamp of special construction containing the rare gas neon. This lamp

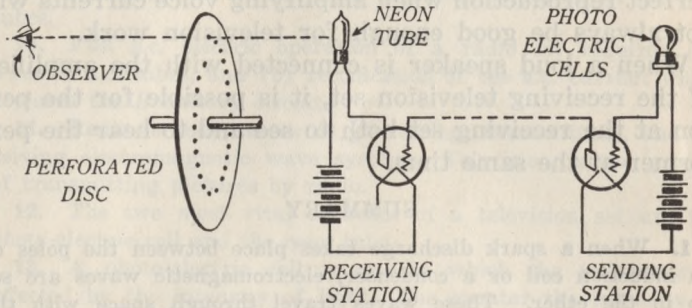


FIG. 462.—Television, receiving set

is of such a nature that in illumination effects it responds with marvelous rapidity to slight changes in current intensity. Every white, gray, or black current impulse imparted by the photo-electric cell to the transmitter of the sending station is picked up by the receiving set, rectified and amplified in the usual manner, and then the picture current is fed to the neon tube. The latter must be capable of responding rapidly and accurately to the fluctuating current; that is, it must vary quickly in its emitted light when the current fed to it varies. The light beam from the neon tube passes through the perforated rotating disc to a screen or directly to the eyes of the observer. The eye of the observer records a light intensity from the neon tube that corresponds with the relative shade of the subject at the sending station. Due to the speed of the disc, the received images follow each other so rapidly that the observer sees the image of the entire object, seemingly all at the same time, although

actually in a series of small portions of the complete object.

The amplifier in a television outfit is a very important factor. If it distorts the picture currents, the resulting received image will be distorted. Amplifiers that give perfect reproduction when amplifying voice currents will not always be good enough for television work.

When a loud speaker is connected with the amplifier of the receiving television set, it is possible for the person at the receiving set both to see and to hear the performer at the same time.

SUMMARY

1. When a spark discharge takes place between the poles of an induction coil or a condenser, electromagnetic waves are set up in the ether. These waves travel through space with the speed of light, that is, 186,000 miles per second = 300,000,000 meters per second.

2. Period T of an electromagnetic wave is the time of one complete vibration; frequency is the number of vibrations per second.

3. $\text{Frequency} = \frac{1}{T} = \frac{velocity}{wave\ length} = \frac{300{,}000{,}000}{wave\ length}$

4. Radio waves are electromagnetic waves of relatively great wave length as compared with light waves. The length of a radio wave is determined by the nature of the mechanism of the transmitter.

5. The conditions necessary for the production of the tones of the human voice are: (a) the production by the oscillator of a continuous series of waves of uniform intensity, and (b) the modulation of these waves by the voice transmitter so as to produce wave pulses of a frequency low enough to be heard by the ear.

6. Radio frequency is the vibration rate of the transmitted wave; audio frequency is the vibration rate of the modulated wave. Radio frequency may be as high as 1,000,000 per second; audio frequency is usually in the neighborhood of 1,000 per second.

7. Two electromagnetic wave systems are in tune when their frequencies are the same, that is, when their periods are equal to each other.

8. The period of a radio wave, and consequently its frequency and wave length, may be varied by changing the inductance L or the capacitance C, the mathematical relation being

$$T = 2\pi\sqrt{LC}$$

9. Amplification is the process of stepping up currents of very low intensity to currents of high intensity by means of electron tubes.

10. Full a.c. electric operation of a radio set requires (a) a set of a.c. tubes, and (b) rectification of the a.c. current which replaces the B battery circuit.

11. Radio is the process of generating, transmitting, and receiving electromagnetic wave systems. Television is the process of transmitting pictures by radio.

12. The two most vital elements of a television set are the photo-electric cell and the neon tube.

13. A photo-electric cell is one in which the resistance is affected by the intensity of light. The greater the intensity of the light falling upon it, the less the resistance of the cell.

14. A neon-tube is an incandescent lamp the illumination of which is strongly affected by very slight changes in current strength.

REVIEW EXERCISES

A

1. Why are ether waves set up by spark discharges called electromagnetic waves?

2. Why are electromagnetic waves sometimes called Hertz or Hertzian waves?

3. What is the velocity of electromagnetic waves?

4. Wherein do electromagnetic waves differ from light waves?

5. What is the relation between period, frequency, wave length, and velocity?

6. Of what do the wireless "dots" and "dashes" consist? What are the vibration limits of audibility of the human ear?

7. What are the essential conditions for radio telephony?

8. What are (a) damped radio waves? (b) continuous waves? (c) modulated waves?

9. What are the necessary conditions for the reproduction of the tones of the human voice by radio?

10. What is the distinction between radio frequency and audio frequency?

11. When are two radio wave systems in tune?

12. Upon what two factors does radio tuning depend?

13. What are the essential parts of a radio receiving set?

14. Make a diagram of a radio tube, and name each part.

B

15. The frequency of a given radio wave is 1000 kilocycles. What is the corresponding wave length?

16. Is it possible for a sending oscillator which is producing 1000 sparks per second to produce radio waves of a greater frequency than 1000?

17. Explain each term in the equation, $T = 2\pi\sqrt{L\,C}$.

18. What is the function of the A battery? the B battery?

19. What is the function of the grid in a radio tube?

20. Explain how the extremely weak currents in a receiving set are amplified.

21. Wherein does an a.c. receiving set differ from a battery set? Is the plate current in an a.c. set charged directly by an alternating current or by a direct current?

C

22. In what particular does a television receiving set differ from the sending set?

23. Describe the scanning disc of a television set, and tell how it works.

24. What is the special property of a photo-electric cell? Where is it used, in the sending set or the receiving set?

25. What is the special property of a neon tube? At which end of the line is it used?

26. What property of human vision is made use of in both moving pictures and television?

CHAPTER XVI

SOUND

The Nature of Sound

481. Introductory. Of the five great divisions of physics, namely, mechanics, heat, electricity, sound, and light, the subject of sound has until quite recently received less attention on the part of investigators than almost any other branch of the science. The reasons for this are manifold, two being especially noteworthy. First, the brilliant discoveries in recent times of the radio-active properties of matter, and the consequent stimulation of research in the related subjects of heat, light, and electricity, have claimed the attention of investigators in these fields; and in the second place, there is no doubt that for many years engineers and others interested directly and indirectly in the physical sciences have felt that while the subject of sound had an academic interest for the physicist and mathematician, it nevertheless was of little practical importance. Happily the day has gone by when we can say just what is and what is not of practical importance. Indeed events within very recent times have taught us that the word "practical" has a much wider application than was formerly dreamed of. Any investigation which widens the bounds of human knowledge or which coördinates scientific data already tabulated may be of the utmost practical importance, as is evidenced, for example, by the application of the principle of photography to the solution of the extremely difficult problems associated with acoustics of large auditoriums.

And, too, there are other and quite different reasons why the subject of sound holds for us, or should hold for us, a most unusual interest. Human life differs in one important respect from other forms of animal life, and that is in its appreciation and application of sound. The wants of the little child are manifested by sounds. Human intercourse is carried on mainly through the medium of sounds—spoken words. And the most aesthetic emotions are aroused by the production of periodic and coördinated sounds, as exemplified in music.

482. Origin of Sound. All sounds originate in vibrating bodies. If a continuous sound be traced to its source, there will always be found a vibrating body. If a tuning fork be set in vibration and then brought in contact with a small glass or ivory ball, the ball will be thrown vigorously from the fork, due to the vibration of the latter. Again, if the prongs of the tuning fork be thrust into water, the latter will be thrown about in a fine spray. Other illustrations of the fact that sound is generated in vibrating bodies may be seen in the vibration of a guitar string. If the string be plucked aside and released, it will give forth a musical note, and at the same time it will appear to spread out into a broad band having a hazy outline, which diminishes to the original size of the string as the sound dies away.

It has been stated that all sounds originate in vibrating bodies. Now bodies may vibrate in a great many different ways. For example, the branches of a tree or the heads of grain in a field vibrate back and forth in the wind, each part in its to and fro motion traveling through the arc of a circle. A chip on the surface of deep water not only moves up and down, but back and forth as well, its motion being elliptical. Also, if a ball attached to a rubber band in a vertical position be pulled downward and then released, it will vibrate up and down in a

straight line. A vibration somewhat similar to this occurs when a glass rod is stroked with a damp cloth. Grasp the rod in the middle with one hand and with a damp cloth in the other hand stroke the rod lightly, between the middle and the end. The glass rod will emit a distinct musical note, and at the same time a tremulous motion will be felt by the hand grasping it. Vibrations run from one end of the rod to the other in a manner somewhat analogous to the up and down motion of the ball attached to the rubber band.

483. Wave Motion. When a disturbance occurs in an elastic medium, such as air or water, waves are set up. A wave motion represents a continuous handing on from



FIG. 463.—Water waves

particle to particle of a disturbance in a medium without an actual transfer of the medium itself. One of the most familiar examples of wave motion is that which occurs on the surface of water. Suppose that a stone be dropped into a lake or a pond, Fig. 463; the waves run out in concentric rings from the point of disturbance. To the observer it would appear that the water is actually being carried forward. This is not true, however, as may be seen by watching the motion of a chip floating upon the surface. As the waves move forward, the motion of the particle upon the surface is up and down. Another excel-

lent illustration of wave motion is seen in the passage of a wind wave over a field of grain. As the wave runs forward each individual head of grain swings back and forth.

The distinction between the motion of a particle of the medium and the motion of the wave itself is of fundamental importance. The bobbing up and down of the chip on the surface of the water and the swinging back and forth of the head of grain are both illustrations of the motion of the particle of the medium. In the case of the water the motion of the particle (the chip) is at *right angles* to the direction of the motion of the waves; in the case of the grain the motion of the particle (head of grain) is back and forth in the *same general direction* as the motion of the wave.

484. Kinds of Wave Motion. There are two kinds of wave motion, transverse and longitudinal.

A *transverse wave* is one in which the vibrating particles move at right angles to the direction of the motion of the wave. A good illustration of transverse waves

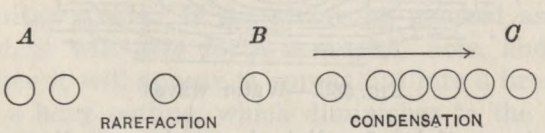


FIG. 464.—Rarefaction and condensation in a longitudinal wave

are those seen on the surface of water; also a series of waves traveling along a rope or string.

A *longitudinal wave* is one in which the vibrating particles move back and forth in the same direction as the motion of the wave. Longitudinal waves may be illustrated by a row of balls suspended as shown in Fig. 464. If a disturbance be set up by striking ball *A*, it will be passed from ball to ball until the disturbance has run the length of the line. This illustrates a longitudinal

wave, which consists of a condensation and rarefaction. As the wave runs forward from A to C there is a point where the particles are crowded together; this is called a *condensation.* At the point B the particles have begun to swing back toward A; this portion of the wave is called a *rarefaction.* A condensation and a rarefaction together constitute the entire wave.

Transverse and longitudinal vibrations may be shown experimentally by means of a metal rod which is clamped at the middle, as shown in Fig. 465. For example, if a steel rod, clamped firmly at the middle, be plucked vertically at one end, it will vibrate transversely; that is, the motion of the rod is at right angles to its length.

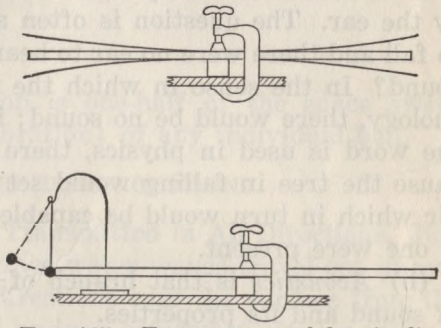


FIG. 465.—Transverse and longitudinal vibrations

The same rod may also be made to transmit longitudinal waves. If it be stroked lightly about one-fourth the distance from the end with a soft piece of leather upon which there has been dusted some powdered resin, the rod will give forth a distinct note of high pitch, and a small suspended ivory ball touching the rod lightly at one end will be thrown out vigorously, Fig. 465, lower. Longitudinal waves, consisting of condensations and rarefactions, run the length of the rod from one end to the other and back again, as shown by the motion of the ball.

Fluids (air and water, for example) are capable of transmitting only longitudinal waves within the medium, that is, waves consisting of condensations and rarefac-

tions. Sound waves in air or in water, then, are transmitted by longitudinal waves, consisting of condensations and rarefactions. It should be noted here, however, that while a wave *in* water is an example of a longitudinal wave, a wave *on the surface* of the water is an example of a transverse wave.

485. Definitions. (a) The word *sound* is used in two distinct senses. From the viewpoint of the psychologist, sound is a sensation; from that of the physicist, sound is that form of vibratory motion which may be perceived by the ear. The question is often asked, If a tree were to fall and there were no ear to hear, would there be any sound? In the sense in which the word is used in psychology, there would be no sound; in the sense in which the word is used in physics, there would be sound, because the tree in falling would set up vibrations of the air which in turn would be capable of affecting the ear, if one were present.

(b) *Acoustics* is that branch of physics which treats of sound and its properties.

(c) *Wave length* is the distance measured in a straight line from any point in a given wave to the corresponding point in the next wave. In transverse waves we usually measure the wave length from crest to crest or from trough to trough, as from *A* to *B*, Fig. 466. In longitudinal waves we measure the wave length from condensation to condensation or from rarefaction to rarefaction. A wave length, however, is not necessarily measured from crest to crest or from condensation to condensation; it may be measured from any point in a wave to the corresponding point in the next wave.

Amplitude of vibration is one-half the distance through which a particle swings as the wave runs under it. In transverse waves the amplitude is measured at right angles to the direction of propagation of the wave,

ac and *de*, Fig. 466. In longitudinal waves, amplitude is measured in the same line as the direction of propagation of the wave. In the case of the swinging balls, the

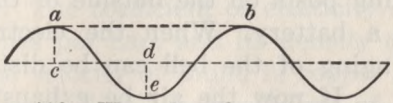


FIG. 466.—Wave length and amplitude

amplitude of vibration is one-half of the space swept out by the swing (vibration) of any individual ball.

TRANSMISSION OF SOUND

486. Sound Waves Transmitted in All Directions. Just as the transverse surface water waves travel out in concentric circles in all directions over the surface from the point of disturbance, so in a somewhat similar manner sound waves in air travel outward in the form of condensations and rarefactions from the point of disturbance in a series of concentric spherical shells. Suppose, for example, a bell is struck. As it vibrates it sets up a series of condensations and rarefactions which travel outward, as shown in Fig. 467. These waves striking upon the ear produce, due to the change of pressure caused by the condensations and rarefactions, a disturbance in the or-



FIG. 467

gans of hearing which give rise to the sensation of hearing.

Sound waves are not transmitted through a vacuum. This may be demonstrated as follows: Place an electric bell within a bell jar, as shown in Fig. 468. Connect the binding posts on the outside of the bell jar to the poles of a battery. When the electric circuit is closed, the ringing of the bell can be distinctly heard. If now the air be exhausted from the bell jar, the sound of the bell becomes fainter and fainter, and finally dies away altogether.

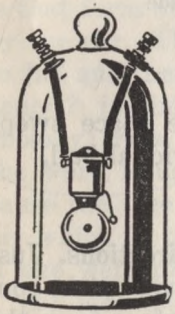


FIG. 468

The characteristics of a medium suitable for the transmission of sound waves are as follows: The medium must be *elastic, continuous,* and *ponderable.* When we say that a medium is ponderable, we mean that it has weight.

When energy from a given source is transmitted in all directions outward through a medium by means of waves (sound waves or light waves, for example), and we think of the waves as moving outward in all directions from the source, we properly speak of the *speed* of transmission, since no definite direction is specified; when, on the other hand, we think of a given portion of a wave front as moving forward in a definite direction, we use the term *velocity.*

487. Velocity of Sound. The velocity of sound in any medium depends upon two factors, (a) the elasticity of the medium, and (b) its density. The greater the coefficient of elasticity e of the medium, the greater the velocity; also, the greater the density d of the medium, the less the velocity. The relation of the velocity of sound

to these two factors is expressed definitely by the equation

$$\text{velocity} = \sqrt{\text{coefficient of elasticity/density}}$$

$$v = \sqrt{\frac{e}{d}}$$

Example 1. The coefficient of elasticity e of copper is 12×10^{11} dynes per cm^2, and its density d is 8.8 g per cm^3. Find the velocity of sound in copper in meters per second. *Solution*: Since the coefficient of elasticity and the density are given in terms of *centimeters,* our equation will give v in cm/sec.

$$v = \sqrt{\frac{e}{d}} = \sqrt{\frac{12 \times 10^{11}}{8.8}} = 369{,}300 \text{ cm/sec.}$$

488. Velocity of Sound in Different Media. The velocity of sound in air was first determined by two observers who stationed themselves several miles apart. A cannon was fired by one observer and the other noted the flash and counted the number of seconds required for the sound to reach him, it being assumed that the time required for the light to travel from the cannon to the observer was so small as to be negligible. In this way a calculation of the velocity of sound in air was made. Many later experiments have been devised and carried out to determine accurately the velocity of sound in different media.

The velocity of sound in air at 0° *C is* 1090 *feet per second, or* 332 *meters per second.*

The velocity of sound in water at 4° C is 4674 feet per second; *that is, the velocity in water is about* 4 *times that in air.*

The velocity of sound in steel is 16,500 feet per second, *or about* 14 *times that of sound in air.*

The reason the velocity of sound in liquids and solids is greater than in air is due to the fact that the co-

efficient of elasticity of liquids and of solids is many times greater than that of air.

489. Relation of Temperature to Velocity. An increase in temperature causes an increase in the velocity of sound in air. The reason for this will be made clear when we consider the equation $v = \sqrt{e/d}$. When air, which is free to expand, is heated, its elastic properties remain unchanged, while its *density is diminished,* and thereby the value of v is increased. On the other hand, if the air be confined in a vessel of approximately constant volume and heated, the air cannot expand, hence there will be no change in its density; the coefficient of elasticity, however, will be *increased,* and as a result the value of v will again be increased. Thus we see that a change in temperature causes an increase in the velocity of sound in air.

A change in temperature of 1° *C above or below* 0° *C causes a corresponding change in velocity of about* 2 *feet per second, or* 0.6 *meter per second.*

EXERCISES. **1.** What is the velocity of sound in air in feet and meters per second when the temperature is +25° C?

2. What is the velocity of sound in air in feet and meters when the temperature is —20° C?

REFLECTION AND REFRACTION OF SOUND

490. Reflection of Sound. When sound waves strike against an obstructing medium, such as the face of a cliff, the wall of a building, the trees of a forest, or even against the vapors of a cloud, they are reflected; that is, they bound off, as does a ball when thrown against a wall. Sound waves reflected from a smooth concave surface may be brought to a focus by a similar concave surface as shown in Fig. 469.

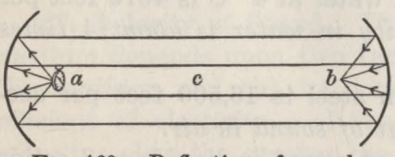


FIG. 469.—Reflection of sound

The ticking of a watch placed at *a* may be distinctly heard at *b*, although it may be entirely inaudible at *c*, midway between the two.

Sound may be reflected from a layer of air having a different density from that in which the sound wave is traveling. For example, a flame or the heated air rising from it makes a fairly good reflecting surface. Also, warm air rising from the earth, due to the heat of the sun, may reflect sound. This explains why given sounds sometimes do not appear to be as intense during the daytime as at night. In the daytime, even when air is still and clear, sound waves traveling from a given source may encounter reflecting layers of air due to the fact of differences of temperature; at night, on the other hand, the air is of a more uniform temperature, and there is therefore less reflection and consequently less loss of energy. Travelers relate that in the long quiet nights of the polar regions sound may be heard for incredible distances.

491. The Echo. If a person calls at a distance from a reflecting medium, in a very short time his voice will be returned to him in an echo. The *echo* is due to the fact that the sound waves traveling outward strike against the medium and are reflected back to the source. The echo is never as loud as the original sound, because a portion of the energy of the wave is always lost by absorption at the reflecting surface, and a small portion is also lost as heat in traveling from the person to the reflecting medium and back again. Since the sensation of sound lasts for about one-tenth of a second, it follows that in order to hear an echo of one's voice, it is necessary to stand far enough away from the reflected surface so that more than one-tenth of a second will elapse between the origin of the sound and its return. Since at ordinary temperature (20° C) sound travels at the rate of about 1100

feet per second, it will therefore be necessary to stand at a distance greater than 55 feet from the reflecting surface in order to hear the echo.

In rooms of ordinary size the walls are so close together that the reflected waves return before the effect of the original sound on the ear has died out. The echo thus blends with and strengthens the original sound instead of interfering with it. This is why, in general, a speaker may be heard so much better indoors than in the open air.

Echoes may sometimes be heard when there is apparently no reflecting surface present. For example, the sound of a whistle or that due to the firing of a gun on the water is often heard to roll away (echo and re-echo) apparently to a great distance. The explanation of this phenomenon lies in the fact that the air, which is optically transparent, is acoustically opaque. The rolling sound which results from a series of echoes is due to reflections from layers of air which are very nearly acoustically impenetrable, and which therefore act as reflecting surfaces.

492. Refraction of Sound. Sound waves may be refracted. *Refraction* is the change of direction of the motion of a wave due to its passage from a medium of one density to a medium of different density. For example, if a series of sound waves originating at the point *A*, Fig. 470, strike against the balloon shaped body *B*, containing carbon dioxide, which is more dense than air, the waves will be bent in such a way that the sound will come to a focus at *C*. This may be readily understood when we

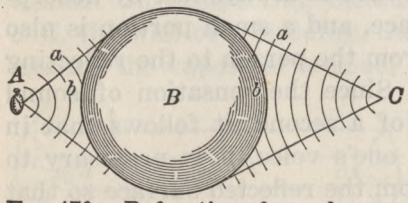


FIG. 470.—Refraction of sound waves

consider what happens to the portion of the wave marked *ab*. Since *b* strikes the dense medium first, its speed is retarded and hence the line of direction is bent downward as shown; *a* gets out of the medium first, and hence travels with a greater speed than *b*, thus bending the wave further towards *C*. Of course, when both portions of the wave get into the air, they travel with the same speed. If the balloon shaped vessel were filled with a gas lighter than air, such as hydrogen, the waves, instead of converging and coming to a focus as shown, would *diverge*.

In the passage through successive layers of air having different densities, sound waves may rise in one region and descend in another, thus producing "islands

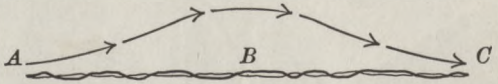


FIG. 471.—Sound heard distinctly at *A* and *C*; silence at *B*

of silence." This condition is very well illustrated in Fig. 471 at *C*. In the regions *A* and *C* the sound may be heard distinctly, while at *B* there is a region of comparative silence. This phenomenon was discovered in 1860 by Joseph Henry who was at that time employed by the Government Light House Bureau. Henry noted that the sound of a steam siren at *A* often passed completely above a vessel at *B*, but could be heard distinctly at a second vessel stationed at *C*. Refraction of this sort is caused apparently by successive layers of air having different densities due to differences in temperature.

ARCHITECTURAL ACOUSTICS

493. Reverberations. The confused sound which comes from the prolongation of a series of sounds resulting from reflections is called a *reverberation*. A reverbera-

tion is different from an echo in that in the case of an echo the original sound and the reflected sound (the echo) constitute two distinct impressions; a reverberation, on the other hand, is a prolonged confusion of sounds resulting from repeated reflections. For ordinary room temperature, an echo requires that the reflecting surface (wall or ceiling) shall be at least 55 feet from the source. Reverberations may arise, and indeed do

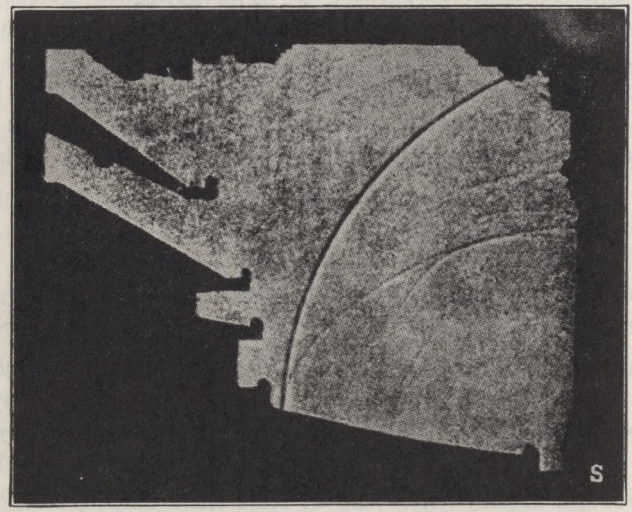


FIG. 472.—Position of a sound wave in an auditorium 0.05 second after leaving the origin at *S*

arise, from reflections from surfaces which are much less than 55 feet from the source, as is illustrated by the confused sounds resulting from reflections which sometimes occur in auditoriums.

It is now possible to photograph sound waves. Some very beautiful photographs of this sort have been obtained by Professor Sabine and applied by him to the practical problem of auditorium acoustics. Two such

photographs are shown in Figs. 472 and 473, in which a vertical section of a theater auditorium is represented. In Fig. 472 a sound wave (heavy curved line) is shown 0.05 second after the sound was produced on the stage at *S*. In Fig. 473 there is shown the position of the reflected waves 0.14 second after the production of the sound. The reflected waves, which seem to come from many directions, were generated by the original impulse at *S*. These

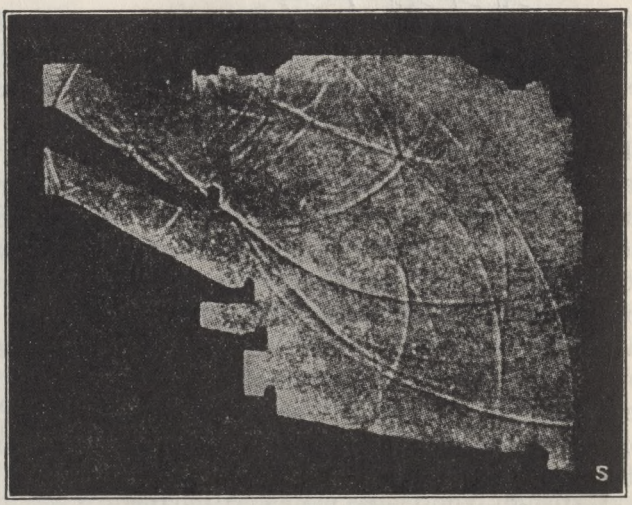


FIG. 473.—Reflected sound waves 0.14 second after the production of the original sound at *S*

multiple waves continue to develop with increasing confusion, until the sound is diffused throughout the entire auditorium, producing the condition called reverberation.

A striking example of reverberation is that which occurs in the Baptistry of Pisa, a building circular in form and having a hemispherical roof. The effect, so far as the reverberations are concerned, is made more conspic-

uous by the good reflecting surface furnished by the polished marble with which the interior of the building is constructed. A single sound originating in the body of this building continues to be audible as a reverberation for 12 to 15 seconds.

A scientific study of the relation of reverberations to the acoustical properties of auditoriums has, within the past few years, led to important discoveries in the domain of architectural acoustics.

494. **Acoustical Properties.** Until quite recent times the science of acoustics as applied to buildings received but scant attention. Not only was this phase of acoustics supposed to involve insurmountable difficulties but the whole subject was invested with an atmosphere of uncertainty and mystery. Many erroneous theories were in vogue and accepted almost without question. As an instance, it was commonly supposed that strings or wires stretched across an auditorium having bad acoustical properties would remedy the defects. We now know that such devices are useless. And further, it was currently reported that two buildings which are exactly alike may have acoustical properties which are quite unlike. This theory, however, turned out to be only half true, based as it was on comparisons of buildings which were only apparently alike, but which in reality differed very materially in important details, especially as relating to the materials which entered into the construction of the interior of the building and the furnishings. We now know that there is no reason why two auditoriums which are alike in all particulars should not have practically the same acoustical properties.

In order that an auditorium may possess satisfactory acoustical properties, it is necessary (a) that the sound reaching each individual member of the audience shall be sufficiently loud; (b) that the various components of

a complex sound shall maintain their proper relative intensities; and (c) that the successive sounds in rapidly shifting articulation, as in speaking or in the production of music, shall be clear and distinct.

495. Acoustical Defects in Auditoriums. The usual acoustical defects in auditoriums are three, namely, echo, dead points, and reverberation.

Echoes arise as a result of the reflection of sound from smooth walls or ceilings. Echo may in general be avoided by breaking up the reflecting surfaces by the use of beams, depressions, and other irregularities. Generally speaking auditoriums are less likely to exhibit troublesome echoes when their outlines are rectangular.

Dead points in an auditorium are usually caused by reflection of sound from curved walls or ceiling. When a condensation of one sound wave coincides with a rarefaction of another the two neutralize each other and a dead point results. Sometimes, however, a dead point may result from a flat reflecting surface. Sabine mentions the case of a theater in which this defect was traced to a flat oval panel in the ceiling. The trouble was corrected by hanging a cloth canopy just below the panel.

Reverberation, as already explained, is a confused sound resulting from repeated reflections. Most acoustical defects in auditoriums are caused by reverberations. A room in which there is too much reverberation is acoustically bad because of the resulting confusion of sounds; and on the other hand, a room in which there is too little reverberation may, musically speaking, be too "dead" to give satisfactory results.

Since reverberation defects are due to reflection, the remedy lies in modifying the reflecting surfaces. Reverberations may, in general, be decreased by covering a part of the walls with some soft material which pos-

sesses good absorbing and poor reflecting properties. An open window may be regarded as a perfect absorber, since all the sound falling upon it passes out.

The *period* or *time of reverberation T* is the number of seconds required for reverberation to die out.

The *coefficient of absorption* of a given substance is 1 divided by the number of square feet of the substance required to produce the same absorbing effect as one square foot of open window. Thus, it was found that about 40 square feet of brick wall were required to produce the same period of reverberation as is produced by one square foot of open window. The coefficient of absorption of brick wall is therefore $1/40 = 0.025$. This means that if the unit of absorption is taken as that of one square foot of open window, the coefficient of absorption of brick wall is numerically equal to 0.025 unit per square foot. The following is a list of the coefficients of absorption of different absorbing materials which are used in the construction and furnishing of auditoriums.

Absorber	*Coefficient*
Open window	1.00 unit per square foot
Hair felt, 1 inch thick	.55 " " " "
Stage opening	.25 " " " "
Curtains	.23 " " " "
Carpets	.20 " " " "
Wood	.061 " " " "
Plaster on lath	.033 " " " "
Cork floor tile	.030 " " " "
Brick wall	.025 " " " "
Audience	4.7 units per person
Seats, upholstered	2.0 " " "
Seats, plain	0.1 " " "

The *absorbing power* of a given area is the product of the area in question and its coefficient of absorption. Since the interior of every hall is made up of various

substances, to get the total absorbing power A we shall have to multiply the area of each substance by its coefficient of absorption and then get the sum of the products thus obtained.

Example. The volume of a given auditorium is 100,000 cu. ft. The absorbing areas are as follows: floor area, covered with cork tile, 4550 sq. ft.; wall area, brick, 2860 sq. ft.; stage opening, 600 sq. ft.; 500 plain wooden seats. Find the total absorption of this auditorium when empty. *Solution*: Referring to the areas given above and the coefficients of absorption given in the table, we have as approximate values

$$\begin{aligned}
\text{floor absorption} &= 4550 \times 0.03 = 137 \\
\text{ceiling} &= 4550 \times 0.033 = 152 \\
\text{brick wall} &= 2860 \times 0.025 = 72 \\
\text{stage} &= 600 \times 0.25 = 150 \\
\text{seats} &= 500 \times 0.1 = 50 \\
&\text{Total absorption } A = \overline{561}
\end{aligned}$$

496. Sabine's Equation. The time required for a reverberation to die away after the cessation of the original sound is called *period of decay or the reverberation time,* and is represented by the letter T. For a given note the value of T depends primarily upon the absorbing power of the interior of the auditorium.

In a hall acoustically good for public speaking, the value of T should be about 1 second. For the best musical effects, however, T should be 2 seconds or more, depending on the size of the auditorium. It appears then that a hall which is acoustically good for speaking may, to a certain extent at least, be bad for the best musical effects, and vice versa. The problem of controlling the reverberation time is therefore an important one. Sabine worked out the following equation which is now uni-

versally employed in preventing and correcting acoustical reverberation defects.

$$T = 0.05 \frac{V}{A}$$

in which T is the reverberation time in seconds, V the volume of the room in cubic feet, and A its total absorption.

Experience with a number of auditoriums of acceptable acoustic properties makes possible the formulation of the following values of reverberation times in seconds for rooms of different volumes, for a maximum audience in each case:

V	T
10,000 cubic feet	0.6 to 0.8 seconds
25,000 cubic feet	.8 to 1.1 "
50,000 cubic feet	.9 to 1.2 "
100,000 cubic feet	1.2 to 1.5 "
200,000 cubic feet	1.4 to 1.7 "
400,000 cubic feet	1.7 to 2.0 "
600,000 cubic feet	1.8 to 2.2 "
800,000 cubic feet	1.9 to 2.3 "

Example. Compute the reverberation time T for the auditorium described in Example, Art. 495, for a full audience of 500 persons. *Solution*: The total absorption A for the auditorium empty was found to be 561 absorption units. To get the total absorption of the hall when filled with people we shall have to add to the above value the difference between the absorption due to 500 people, which is $500 \times 4.7 = 2350$, and the absorption due to 500 empty seats, as previously computed, giving us a net increased absorption of $2350 - 50 = 2300$ units. The total absorption for a full audience, then, is $A = 561 + 2300 = 2861$. The reverberation time is

$$T = 0.05 \frac{V}{A} = \frac{0.05 \times 100{,}000}{2861} = 1.7 \text{ seconds}$$

Intensity and Loudness of Sound

497. Distinction between Intensity and Loudness. *Intensity* of sound may be defined as the quantity of energy in the sound wave passing through a surface of unit area (1 cm^2) in one second. It is understood that the unit area is at right angles to the direction of propagation of the sound wave. Intensity of sound depends upon several factors as follows: (a) distance from the origin; (b) amplitude of vibrations; (c) frequency of vibration; (d) density of the medium; (e) area of the vibrating body.

Loudness is a term used to designate the strength of the sensation received through the medium of the ear. The loudness with which a given sound is heard depends in part on the energy of the vibrating system and in part on the nature of the hearing mechanism of the individual. It is probable that no two persons hear the same sound with exactly the same degree of loudness; indeed, the loudness with which a given individual perceives a sound varies from time to time, depending upon the state of health and nervous condition.

Intensity, being a physical quantity, can be measured. Loudness, on the other hand, cannot be measured in physical units. Loudness is not a physical quantity; it is a sensation. We may say, however, that within certain limits *loudness is proportional to intensity, that is, the greater the intensity the greater the loudness.*

498. Relation of Intensity and Loudness to Distance. It is a common experience that if we walk away from a sounding body the loudness of the sound decreases as the distance increases. Now if the medium be of such a nature that the sound waves move out from the source in a spherical form, and also that there is little loss of energy, due to friction between the particles of the me-

dium, it may be shown both experimentally and mathematically that the *intensity,* and consequently the loudness, *is inversely proportional to the square of the distance from the source.* That is, at a point one mile from the source a person would hear the sound with a given loudness, while at another point two miles from the source the same person would hear the sound only one fourth as loud.

It should be noted, however, that this exact relationship between intensity and the square of the distance holds strictly only in the case of waves spreading in a spherical form. Owing to the presence of reflecting surfaces, sound waves produced under ordinary circumstances do not remain spherical and consequently the inverse square law does not apply exactly to sound waves produced on the surface of the earth. Also, in the case of sound waves traveling in a tube (as for example, a speaking tube) the energy is transmitted from layer to layer of equal areas without much loss due to the spreading of the wave, and consequently sound will travel in a confined space, such as a tube, to a relatively great distance without much diminution in intensity. This fact explains the efficiency of the speaking tube and the megaphone.

499. Relation of Intensity and Loudness to Amplitude of Vibration and Frequency. If, for example, the prongs of a tuning fork be tapped lightly so that they vibrate through a very small arc, a faint sound is heard; if now they be struck sharply, so that the prongs vibrate through a very much larger arc, they give rise to a sound much louder than the first. It can be shown that in the case of simple vibratory motions the *intensity of the sound is directly proportional to the square of the amplitude of the vibrating medium.* If a vibrating body having an amplitude of one unit produces a sound of given intensity, and if then the body be made to vibrate through an

amplitude of 2 units, the intensity will be 4 times as great.

Also, it may be shown that for simple vibratory motions the *intensity is directly proportional to the square of the frequency.*

500. Intensity and Loudness Affected by the Density of the Medium and by the Area of the Sounding Body. As one ascends in a balloon the air becomes rarer and rarer as the distance from the earth increases, and it has been observed that the loudness of the sound diminishes. If, on the other hand, a person goes down into a deep mine or down in a diving bell, the sound of the human voice will be much louder than at the surface of the earth. We may therefore conclude that for air *the rarer the medium the less loud is the sound, and the denser the medium the louder is the sound.*

Also, the loudness of sound is affected by the area of the sounding body. This may be shown by the following simple experiment: If a tuning fork be set in vibration and held in the hand, a sound of a given loudness will be heard. Now if the base of the tuning fork, while still in vibration, be placed in contact with the top of the desk or upon a resonance box, the sound will at once become very much louder, due to the fact that a body of increased area is set in vibration.

EXERCISES. **3.** A given sound originates at a point in air under such conditions that simple sound waves travel outward from the source in spherical form. The sound at 100 feet from the source has a given intensity. What is the relative intensity at a point 300 feet from the source?

4. Suppose that the frequency of the sound wave be doubled and the amplitude be made three times as great, how will the intensity at a given point be affected?

RESONANCE AND INTERFERENCE

501. Free and Forced Vibrations. The vibration of a

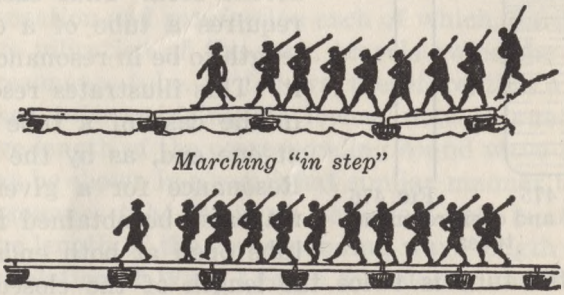
pendulum free from all influence, except that due to the attraction of gravity, is an example of a *free vibration.* If there were no friction, the pendulum would continue to vibrate forever. As it is, however, all freely vibrating bodies sooner or later come to rest, due to the damping effect of the friction of the medium in which they swing. Examples of free vibrations are seen in the motion of the pendulum, as described above, also in the motion of a guitar string, or in that of a tuning fork.

When a body vibrates not in response to its own nature but in response to some periodic force applied to it, we have what is called a *forced vibration.* A good illustration of a forced vibration is that of the motion of the pendulum of a clock or the balance wheel of a watch, in which case the vibrating bodies move in response to a periodic force applied by the weight or the spring. Other examples of forced vibrations are those of the sounding board of a piano, or the body of a violin, or the case of the desk when the stem of the tuning fork is pressed upon it.

502. Resonance. When the period of the vibrating body is the same as that of the impressed force, we have a special case of forced vibration known as *resonance.* If, for example, two tuning forks of the same period be mounted near to each other, and one is set in vibration and after a few moments stopped, it may be observed that the other fork is vibrating, as may be demonstrated by a light ivory ball suspended so as just to touch the fork. This is due to the fact that the impulses of the air from the first fork act upon the second, and since its period of vibration is the same as that of the impulses impressed upon it, it responds. A further illustration of the relation of the impressed force to the period of the vibrating system is seen in the case of a swing. A little child may set a heavy swing in vibration, provided

the impulses (pushes) be so timed as to coincide with the period of vibration of the swing. First the swing moves through a very small arc, but soon it responds to the periodic impulses, and may after a time swing through a large arc. The motion of the swing is an example of a forced vibration, and since its period of vibration coincides with the impulses imparted by the child, we have a case of resonance.

A striking illustration of resonance is that of the vibration of a light bridge, due to the footfalls of soldiers stepping in unison, Fig. 474. If the period of the impulses



Marching "in step"

Marching "out of step"

FIG. 474.—Resonance illustrated by soldiers marching across a light bridge

imparted by the soldiers agrees with the period of the bridge, the latter will be set in vibration. Soldiers when crossing a bridge are therefore usually commanded to break step to avoid the possibility of doing damage to the structure, due to the resonance in response to the regular foot falls.

503. Resonators. A tube so adjusted that the air within it vibrates in unison with some outside vibrating body, such as a tuning fork, is called a *resonator*. If a glass tube, Fig. 475, be partly immersed in water and moved up and down while a fork vibrates above it, there

will be found a certain point at which the tube will respond to the fork, that is, the resonance tube will give out a relatively loud sound. Now if a fork having a different vibration rate be held above the tube, no response is heard. If, however, the tube be moved up or down, there will be found another point at which it will respond to the second fork. Thus each fork requires a tube of a certain length to be in resonance with it. This illustrates resonance in the case of a tube closed at one end, as by the water. Resonance for a given fork may also be obtained from a tube open at both ends, provided the tube is twice the length of the closed tube, Fig. 476.

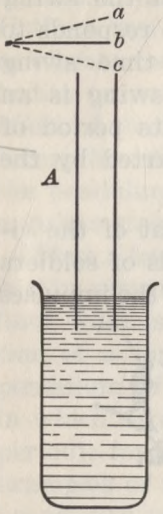


FIG. 475

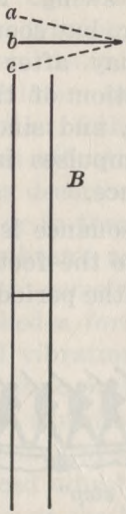


FIG. 476

Closed and open resonance tubes

504. Explanation of Resonance in a Tube. Let *a*, *b*, *c* represent the motion of one prong of a tuning fork which is vibrating at the mouth of a closed resonance tube, so adjusted as to respond to the fork. As the prong of the fork moves from *a* to *b* a condensation is sent down into the tube. While the fork moves from *b* to *c* and comes to rest for a moment in changing its direction of motion, the condensation has time to run down the tube, strike the water, be reflected, and get back to *c* in time to join the new condensation which is formed as the fork moves from *c* toward *b*. The condensations and rarefactions which are reflected from the bottom of the tube coincide with the condensations and rarefactions of

the sound waves given off by the fork, thus producing a re-enforcement of the sound, and hence giving rise to an increase in loudness.

505. Relation of Wave Length to Resonance Tube Length. Let us consider again the case of a resonance tube which is closed at one end and which is responding to a given fork. A sound wave travels one wave length while the fork makes one complete vibration. Now we have learned that during one complete vibration of the fork a condensation travels twice the length of the tube (down and back), and a rarefaction likewise travels twice the length of the tube. A sound wave consists of a condensation and rarefaction each of which, during one complete vibration of the fork, travels twice the length of the resonance tube. It follows, therefore, that a *closed resonance tube* sounding its lowest note is *one-fourth* the wave length of the corresponding sound wave in air.

It may be shown in a somewhat similar manner that an *open resonance tube* sounding its lowest note is *one-half* the wave length of the corresponding wave length in air.

506. Relation of Wave Length to Velocity and Number of Vibrations. Suppose that the fork is making 256 vibrations per second and we desire to find the wave length of the corresponding sound waves in air when the temperature is 20° C. Now the velocity of sound at 20° C is 1130 feet per second. While the fork is making 256 vibrations sound travels a distance of 1130 feet. This means that in a space of 1130 feet there are 256 waves. Therefore, the length of one wave = 1130/256 = 4.4 feet. It follows then that the length of a wave in an elastic medium may be represented by the equation

$$wave\ length = \frac{velocity}{frequency}$$

$$l = \frac{v}{n}$$

in which l is the wave length, v the velocity of the sound in air, and n the frequency, that is, the number of vibrations per second.

507. How to Measure Wave Length and Number of Vibrations by Means of a Resonance Tube. It is possible by means of a resonance tube to measure with a reasonable degree of accuracy the wave length of the sound given off by a tuning fork, and also to determine the number of vibrations per second, that is, its frequency. Suppose that we wish to determine the wave length and the number of vibrations per second of a given tuning fork. A glass tube one end of which is thrust into water, as already explained, serves as a resonance tube. The given fork is set in vibration and held above the tube, which is moved up and down until the point of resonance is reached. We now measure the distance from the fork to the water, which represents the length of the closed resonance tube. Four times this length equals the wave length of the corresponding sound wave in air.

Now by means of the relation, $l = v/n$, we may at once determine n, the number of vibrations per second. Suppose the temperature at which the experiment is carried out is 26° C and that the length of the resonance tube is 10.5 inches. At a temperature of 26° C the velocity of sound, v, in air $= 1090 + (2 \times 26) = 1142$ feet per second; and wave length $l = 4 \times 10.5$ inches $= 42$ inches $= 3.5$ feet. Now using the equation above, we may write

$$n = \frac{v}{l} = \frac{1142}{3.5} = 326 \textit{ vibrations per second}$$

Example. A closed resonance tube 18 inches in length responds to a given fork which makes 188.5 vibrations per second. Find the temperature at the time when the

experiment was performed. *Solution*: Length of tube = 18 in. = 1.5 ft., from which wave length $l = 4 \times 1.5 = 6$ ft. Frequency $n = 188.5$, and velocity $v = 1090 + (2 \times t)$, in which t is the temperature. We may now write

$$l = \frac{v}{n} = 6 = \frac{1090 + (2 \times t)}{188.5}, \text{ from which}$$

$$t = 20.5^\circ \text{ C}$$

EXERCISES. **5.** A closed resonance tube 1 ft. in length responds to a fork making 288 vibrations per second. Find (a) the velocity of sound corresponding to the conditions of the experiment; (b) the temperature.

6. Find the length of a closed pipe that will respond to a fork making 256 vibrations when the temperature is 10° C.

7. Determine the frequency of a fork that is in resonance with a closed tube 15 in. in length at 0° C.

508. Re-enforcement and Interference of Sound. If two sound waves occur together so that the condensations and rarefactions of one coincide with the condensations and rarefactions of the other, the resulting sound will be louder than either of the sounds producing it. In Fig. 477 there is shown in a graphic way the result of adding two sound waves, as has just been described. The resultant sound wave is represented by the line having an amplitude greater than either of the other waves.

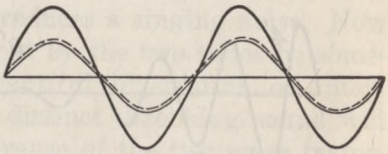


FIG. 477.—Re-enforcement of sound

On the other hand, if two sound waves be superimposed one upon the other, so that the condensation of one coincides with the rarefaction of the other, the result will be a diminution in loudness. Indeed it is possible for two sound waves to be so impressed one upon the other that silence will result. This condition is shown in Fig.

478, in which is represented graphically two sound waves of the same wave length and amplitude, the condensations of one coinciding exactly with the rarefactions of the other. The resultant is represented by the straight line, which indicates silence.

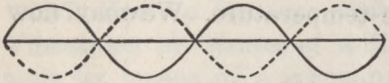


FIG. 478.—Interference of two sound waves resulting in silence

The amplitude of the resultant sound wave is at every point equal to the algebraic sum of the amplitudes of the original sound waves producing it.

509. Beats. Suppose that we have two sound waves of different wave lengths traveling in the same direction, as shown in Fig. 479. At points A and C the waves coincide in such a way that they re-enforce each other. At B they interfere in such a way as partially to annul each

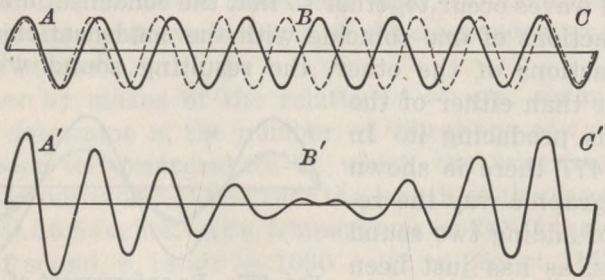


FIG. 479.—Interference of two sound waves producing beats

other. The resultant wave is represented by the line $A'B'C'$. At the points A' and C' the sound will be loud; at the point B' it will be faint, thus producing a rise and fall in loudness. This periodic increase and decrease in loudness due to the interference of two sound waves is called *beats*.

A striking illustration of beats occurs in connection

with the singing flame. Let two jets of flame be prepared as shown in Figs. 480-481. First, over one of the flames, Fig. 480, place a glass tube about 2 feet in length and from three-quarters to an inch in diameter. Move the tube up and down until a certain position is found in which the flame will sing, due to the sympathetic vibrations set up in the tube by the fluttering of the flame. Now over the second flame, Fig. 481, place a similar tube, having around one end a cylinder of paper which may be moved up or down as the case may require. The paper slider may be adjusted until the point is reached at which the second tube produces a singing noise. Now if the wave lengths given out by the two tubes be somewhat different, a very beautiful illustration of interference (beats) occurs. A distinct throbbing sound will be heard due to the interference of the two wave trains.

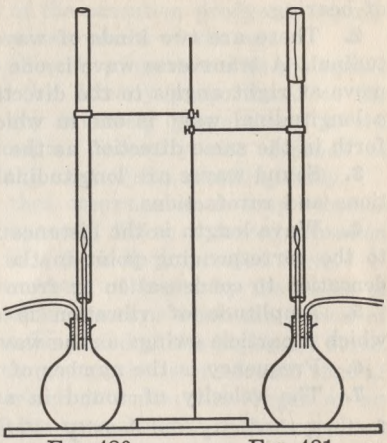


FIG. 480 FIG. 481
Production of beats by the singing flame

The number of beats which occur per second is equal to the difference between the frequencies of the two sounding bodies. For example, if two forks *A* and *B* are sounding together, the frequency of *A* being 256 and that of *B* 292, the number of beats occurring per second will be $292 - 256 = 36$.

SUMMARY

1. All sounds originate in vibrating bodies. Considered in a psychological sense, sound is a sensation; considered in a physical

sense, sound is a vibratory motion capable of affecting the sense of hearing.

2. There are two kinds of wave motion, transverse and longitudinal. A transverse wave is one in which the vibrating particles move at right angles to the direction of the motion of the wave; a longitudinal wave is one in which the particles move back and forth in the same direction as the motion of the wave.

3. Sound waves are longitudinal waves, consisting of condensations and rarefactions.

4. Wave length is the distance from any point in a given wave to the corresponding point in the next wave, that is, from condensation to condensation or from rarefaction to rarefaction.

5. Amplitude of vibration is one-half the distance through which a particle swings as the wave runs under it.

6. Frequency is the number of vibrations per second.

7. The velocity of sound in any medium depends upon two factors, elasticity and density. $Velocity = \sqrt{e/d}$.

8. The velocity of sound in air at 0° C is 1090 feet per second. A change of temperature of 1° C above or below 0° C produces a corresponding change (increase or decrease) of about 2 feet or 0.6 meter per second.

9. An echo is the repetition of a sound by reflection from some distant surface. In order to hear an echo one must be 55 feet or more from the reflecting surface.

10. A reverberation is a confused sound resulting from a number of reflections.

11. Acoustical defects in auditoriums are in general due to one or more of the following causes: (a) echoes, (b) dead points, (c) reverberations. Reverberation is the main cause of acoustical defects.

12. The period or time T of reverberation is the number of seconds required for a reverberation to die out.

13. Sabine's law: $T = 0.05 \times V/A$, in which T is the time of reverberation, V the volume of the room in cubic feet, and A the total absorption.

14. The coefficient of absorption for any given substance is 1 divided by the number of square feet of the substance required to produce the same absorption effect as that of 1 square foot of open window.

15. Intensity of sound is the quantity of energy in the sound wave which passes through unit area (1 cm^2) per second. Loud-

ness of a sound is the strength of the sensation produced, depending upon the intensity and the sensitiveness of the ear.

16. Intensity of sound is (a) directly proportional to the square of the amplitude of vibration, (b) directly proportional to the square of the frequency, and (c) inversely proportional to the square of distance from the origin. Intensity also depends on the density of the medium and the area of the sounding body.

17. A free vibration is one that occurs when a body vibrates independently of any periodic impressed force; a forced vibration is one that occurs in response to a periodic impressed force.

18. Resonance is the re-enforcement of sound by the union of direct and reflected waves. Resonance occurs when the period of the vibrating body is the same as the period of the impressed force.

19. The length of a closed resonance tube is one-fourth that of the wave length of the sound which it emits; the length of an open resonance tube is one-half the wave length of the sound.

20. The relation between the wave length l of a sound, the velocity v, and the frequency n may be expressed as follows:

$$l = v/n, \text{ from which}$$
$$n = v/l, \text{ and}$$
$$v = nl$$

21. Two sound waves (a) may re-enforce each other so as to produce a sound of greater intensity, (b) may interfere so as to produce a sound of less intensity, or (c) may produce beats.

22. Beats are periodic changes (rise and fall) in the intensity of a sound due to the interference of two or more sound waves of different wave lengths.

23. The number of beats produced by two sounding bodies is equal to the difference between their frequencies; that is, number of beats $= n - n'$.

REVIEW EXERCISES

A

1. What is the difference between a transverse wave and a longitudinal wave?

2. Explain the terms condensation and rarefaction, as used in connection with longitudinal waves.

3. Compare the direction of the motion of a wave with the motion of a vibrating particle in the case of (a) surface water wave;

(b) a wave passing over a field of grain; (c) a sound wave in air.

4. What is the distinction between the psychological and physical definitions of sound? What is acoustics?

5. Define and illustrate by a sketch (a) wave length, (b) amplitude of vibration.

6. Upon what two factors does the velocity of sound depend? Explain the meaning of each term in the equation $v = \sqrt{e/d}$.

7. Why does the timer in a 200 yard dash start his stop watch by the flash of the pistol rather than by its report?

8. What is the velocity of sound in air, in feet, at (a) 0° C? (b) —20° C? (c) +20° C?

9. What is the approximate relation of the velocity of sound in air to that in water? iron?

10. A flash of lightning is seen and 5 seconds later the thunder is heard. How far away is the lightning discharge, the temperature being 25° C?

11. Why must a person be 55 feet or more from a reflecting surface in order to hear an echo?

12. A sound is emitted from a submarine, and 4 seconds later the echo from the bottom of the ocean is detected. How far above the bottom is the submarine, the velocity of sound in sea water being 4800 ft. per second?

13. The pilot on a steamer blows a whistle and 8 seconds later hears an echo from a distant iceberg. How far away is the iceberg, the temperature being 50° F?

14. What is the difference between a reverberation and an echo?

15. What are the three main causes of acoustical defects in auditoriums? How may each be remedied?

16. What is refraction of sound? Explain why sound waves come to a focus when falling upon a spherical shaped balloon containing a gas which is denser than air.

17. A balloon shaped vessel is filled with hydrogen. What will happen to sound waves which pass from air through this vessel? Explain.

18. How do you account for the "islands of silence" that sometimes occur on the ocean?

19. What is the distinction between intensity and loudness of sound? Upon what factors does intensity depend?

20. What is the relation of intensity to (a) distance? (b) amplitude of vibration?

21. Explain why striking a bell a vigorous blow causes it to give off a louder sound than if it is tapped lightly.

22. Compare the intensity of a given sound at two points, one a half mile from the source, the other three miles from the source.

23. An electric bell is ringing in a bell jar which is connected to a vacuum pump. What happens to the sound of the bell as the air is exhausted from the jar? Explain.

24. The stem of a tuning fork which is in vibration is placed in contact with the top of a table. What change in the sound occurs? Explain.

25. Two men, *A* and *B*, are 1000 and 2000 feet respectively from a bell. How many times as loud does the bell sound to *A* as to *B*?

26. What is resonance? Describe an experiment to illustrate resonance of sound.

27. Under what conditions may two sound waves interfere so as to produce silence?

28. What is the relation between wave length of sound, velocity, and frequency? Write the equation.

29. What is the relation between the wave length of sound in air and the length of (a) a closed resonance tube? (b) an open resonance tube?

30. A closed resonance tube 2 ft. in length responds to a given fork when the temperature is 15° C. Find the frequency of the fork.

B

31. A man sets his watch at noon by the sound of a factory whistle which is 3 miles away. The temperature of the air is 20° C. How many seconds slow will his watch be as compared with the factory time?

32. Five square feet of a certain material on the floor of an auditorium has the same absorbing effect as 1 sq. ft. of open window. (a) What is the coefficient of absorption of this material? (b) What is the material?

33. Upon what two factors does the period of reverberation of a room depend? State Sabine's law.

34. In a certain auditorium having a volume of 100,000 cu. ft. it was found that the reverberation time was 1.3 seconds. Are the

acoustical properties of this auditorium good or bad? Explain.

35. Classify the following as free or forced vibrations: (a) the vibration of a pendulum swinging under the influence of gravity alone; (b) the vibration of a tuning fork, (c) the pendulum of a clock, (d) the sounding board of a piano, (e) the air in a resonance tube above which there is vibrating a tuning fork.

36. A closed resonance tube 18 in. in length responds to a fork which makes 190 vibrations per second. Find the temperature.

37. A closed organ pipe is 3 ft. long. (a) What is the wave length of its fundamental (lowest) tone? (b) How long must an open pipe be to give a tone of the same wave length, the temperature remaining the same?

38. What are beats, and how are they produced?

39. Two tuning forks, *A* and *B*, when sounding together produce 40 beats per second. (a) What is the difference between the frequencies of the forks? (b) If *A* makes 256 vibrations per second, what may be the frequency of *B*?

C

40. Explain why an increase in temperature causes an increase in the velocity of sound in air.

41. Explain why the velocity of sound in metals is greater than in air.

42. Two men, *A* and *B*, on the plains are 4 miles apart. The wind is blowing 10.5 miles per hour directly from *A* to *B*. A gun is fired at a point midway between the men at a time when the velocity of sound is 1100 feet per second. How much sooner will *B* hear the report of the gun than *A*?

43. A canyon running in a north-south direction is a mile wide. On a day when the temperature is 20° C, a man in the canyon fires a shot and 3 seconds later hears an echo from the east wall of the canyon. How far from the east wall is the man?

44. How many seconds after firing the shot (Ex. 43) does he hear the echo from the west wall?

45. Make a drawing to illustrate the effect of sound waves upon each other such that they produce (a) a louder sound, (b) silence, (c) beats.

CHAPTER XVII

SOUND (Continued)

PITCH AND MUSIC

510. Pitch. The *pitch* of a sound is that characteristic which is determined by its frequency. When the number of vibrations per second is great, the pitch is high; when the number of vibrations is small, the pitch is low. If two sounds are produced by the same number of vibrations per second, they are said to have the same pitch.

If a tuning fork is tapped lightly, it gives off a low sound of a given pitch. If now the fork be struck a sharp blow, it gives off a loud sound, but of the same pitch as before. The sound from a vibrating body may be faint or loud depending on the amplitude of vibration, but so long as its vibration rate (frequency) remains the same, its pitch does not change.

A case of special interest is furnished by a locomotive which is sounding its whistle while running at high speed. *The pitch of the approaching whistle is higher than when the train is standing still; and the pitch of the retreating whistle, as the train recedes from the observer, is lower than it is when the train is at rest.* That is to say, when a sounding body moves toward or away from the observer, the pitch is apparently altered. When the train is standing still and the whistle is sounding, the ear receives a given number of sound waves per second. When the train is approaching, the pitch is apparently higher because the ear receives a greater number of waves per second, and when the source of sound retreats, the pitch is apparently lower because the ear

receives a fewer number of waves per second. This apparent change of pitch, due to the motion of a sounding body is known as the *Doppler effect,* so named after Christian Doppler, who worked out the following equations expressing the relation between the change of pitch n' and the velocity V:

(a) When the sounding body is moving *toward* the observer, the apparent pitch is

$$n' = \frac{nV}{V - v}$$

in which n' is the apparent pitch, n the actual pitch, V the velocity of sound, and v the velocity of the moving body.

(b) When the sounding body is moving *away* from the observer,

$$n' = \frac{nV}{V + v}$$

511. The Siren. Pitch may be measured by means of a *siren,* which consists of a circular disc having several concentric rows of holes, and which is mounted so as to rotate on its central axis at a given speed, as shown in Fig. 482. If a stream of air be blown through a tube against the rapidly rotating disc in line with a given row of holes, a note of a definite pitch is produced. When one of the holes of the rotating disc comes in front of the tube, a puff of air goes through, thus producing a condensation; at the next instant, the air is cut off by the disc, producing something corresponding to a rarefac-

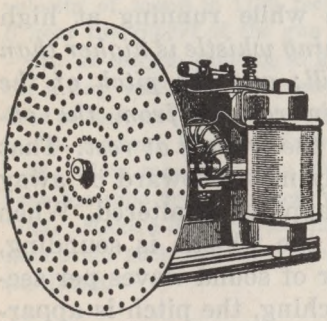


FIG. 482.—Siren, driven by an electric motor

tion. Now if the disc be rotated at a given speed and the tube through which the air is blown be slowly moved from the center of the disc toward the circumference, a series of sounds will be produced which increase in pitch from low to high notes, due to the fact that the holes in the disc increase in number from the axis outward. To determine the pitch of a given note as produced by the siren, we must know the number of holes in the particular series in front of the tube and also the number of the rotations of the disc per second. Suppose that in the outside row there are 48 holes and the disc is making 10 revolutions per second. This will give $10 \times 48 = 480$ condensations and rarefactions, that is, 480 vibrations per second. The resulting note is one having a pitch due to a frequency of 480.

If a siren be provided with an apparatus that will count automatically the number of revolutions per second, it is possible by means of this instrument to determine the number of vibrations per second of a given sound, provided the operator is able to tell by ear when the two sounds, that of the siren and that of the given body, are in unison. Suppose, for example, that we wish to determine the number of vibrations per second made by the wings of a bee buzzing on the window pane. The speed of rotation of the siren is gradually increased until the pitch of the instrument coincides with that of the wings of the insect. At this instant the number of revolutions per second is noted. The number of revolutions per second multiplied into the number of holes passing before the tube gives the number of vibrations required. In a similar manner the frequency of any instrument, such as a tuning fork, may be determined.

512. Musical Sounds. A *musical sound* is one which has regularity of pitch and which produces an agreeable effect upon the ear.

A *noise* is a combination of sounds in which the ear is unable to detect a definite pitch, as in the confused sounds resulting from street traffic or in the clatter of machinery in a factory. Relatively speaking, noises are sounds which produce disagreeable or irritating effects on the hearer. According to this way of putting it, then, it follows that what may be classified as a noise by one person or people may be considered as music by another. A notable case in point was that furnished by Wagner's "Tannhäuser Overture." Soon after the appearance of this famous overture a German critic of Frankfort likened it to a display of "shrill noises and broken crockery effects."

The essential characteristics of musical sounds are (a) regularity of pitch, and (b) agreeableness to the ear. A series of sounds may have a definite pitch and may be periodic and yet not be pleasing to the ear, as in the case of the beating of the tom-tom in an oriental temple, or in the monotonous cries which accompany an Indian war dance. Saxophone music, jazz, distorted radio sounds, and the sounds emitted by most automobile horns may be taken as border-line cases between musical sounds and noise.

513. Some Musical Terms. (a) A *musical tone* is a train of sound waves of constant frequency. A *note,* as used in a musical sense, may refer to a sound having a definite pitch or to a symbol in an arbitrary scale.

(b) A *musical interval* is the ratio of the pitch (vibration number) of a given note to the pitch of another note. For example, the interval expressed by the ratio 1:1 is called *unison*; that expressed by the ratio 2:1, an *octave.* Thus if one note has a pitch due to 256 vibrations per second, and another 512 vibrations per second, we say that the pitch of the second is an octave higher than the first.

(c) When any three notes having ratios 4:5:6 are sounded together, they produce a tone which is pleasing to the ear; the same is true also of the ratios 10:12:15.

A group of three notes having the ratios 4:5:6 is called a *major triad*, and a group having the ratios 4:5:6:8 is a *major chord*, which is represented in music by the notes *do, mi, sol, do'*. In Fig. 483 there is shown a set of four forks which sounding together give a major chord. The major triad is sometimes called a major chord.

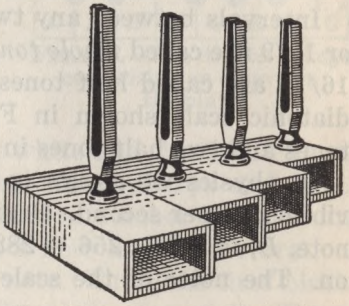


FIG. 483.—Four forks which sounding together give a major chord

The ratios 10:12:16 constitute a *minor* triad, and a group of four notes having the ratios 10:12:16:20 is a *minor chord*. From the major and minor chords we derive our musical scales.

514. The Diatonic Scale. The *diatonic scale* or *gamut*,

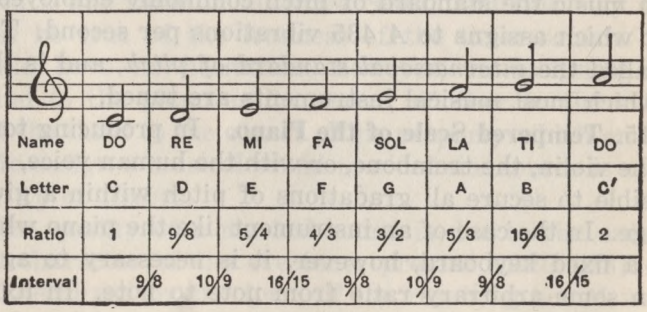


FIG. 484.—Diatonic scale

as it is sometimes called, consists of a series of eight notes having definite ratios. The scale derived from the

major chord (ratios = 4:5:6:8) is called the *major diatonic scale*; that based on the minor chord (10:12:15:20) is called the *minor diatonic scale*. The notes of the major diatonic scale are given in Fig. 484.

Intervals between any two notes having the ratios 9/8 or 10/9 are called *whole tones*; intervals having the ratios 16/15 are called half tones. It will be noted that in the diatonic scale shown in Fig. 484 there are five whole tones and two half tones in an octave from *C* to *C'*.

In physics we assign to middle *C* a frequency of 256 vibrations per second. Beginning with *C* as 256, the next note, *D*, is $9/8 \times 256 = 288$; E, $5/4 \times 256 = 320$, and so on. The notes of the scale in the key of C are then,

Key of C	*C*	*D*	*E*	*F*	*G*	*A*	*B*	*C'*
	256	288	320	341	384	426	480	512

The scale may be extended above or below this octave by again using the ratios 9/8, 5/4, 4/3, etc. For example. the note which is an octave above *C'*, that is *C''*, makes $2 \times 512 = 1024$ vibrations per second; the note an octave below *C* makes $\frac{1}{2} \times 256 = 128$ vibrations per second.

In music the standard of pitch commonly employed is that which assigns to *A* 435 vibrations per second. This is called the *international standard of pitch*, and is that to which most musical instruments are tuned.

515. Tempered Scale of the Piano. In producing tones on the violin, the trombone, or with the human voice, it is possible to secure all gradations of pitch within a given range. In the case of an instrument like the piano which has a fixed keyboard, however, it is necessary to agree upon some arbitrary ratio from note to note. In fixing the ratios from note to note on the piano, musicians have agreed to adopt a system known as *equal temperament*; that is, the ratios between all notes are equal. The ratio selected as being most satisfactory is the twelfth root of

2, which is equal to 1.05946. Thus, if we fix the value of the note *A*, Fig. 485, as 435, then the sharp of *A*, the note of the black key next above it, will be $1.05946 \times 435 = 460.9$; the next key *B* will be $1.05946 \times 460.9 = 488.3$, and so on.

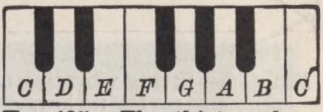


FIG. 485.—The thirteen keys comprising an octave on the piano

The keys representing an octave on a piano are shown in Fig. 485, there being eight white and five black keys. The black keys represent notes called sharps and flats. A *sharp* is a note having a vibration number higher than that of a given note; a *flat* is a note having a vibration number lower than the given note. Thus, the first black key above *C* is the sharp of *C* and the flat of *D*. These thirteen notes, eight white and five black, constitute an octave.

516. Quality of Sound. Sounds may differ in three respects; namely, (a) in loudness, (b) in pitch, (c) in quality. What we mean by *quality* or *timbre* may be illustrated as follows: If middle *C*, for example, be simultaneously struck on a piano, a violin, and a cornet, the pitch is the same in all three cases; and this may likewise be true of the loudness of the three sounds. We would have no hesitation, however, in assigning one sound to the piano as its source, another to the violin, and the third to the cornet. This property which enables us to assign a musical sound to a definite source is called *quality* or *timbre*.

The quality of a musical sound is due to the number and character of the overtones present (Art. 533) and is determined by the form of the resulting sound wave.

VIBRATION OF STRINGS

517. Segmental Vibration of Strings. (a) If one end

of a long flexible rope be fastened to the wall and the other end moved up and down by the hand in such a way that the rope is caused to swing as a whole, in somewhat the same manner as when used by children in "skipping rope," we will have an example of a string vibrating in a single segment, Fig. 486. The rope does not vibrate in a vertical plane, but has a somewhat circular motion; this is true, also, of all vibrating strings. If now the hand be moved more rapidly, the rope may be made to break up into two, three, or more segments, as shown in Fig. 487, depending upon the rate at which the impulses are imparted. The points of least motion, *N* and *N*, are called *nodes*; the points of greatest motion, *A* and and *A*, are called *antinodes*.

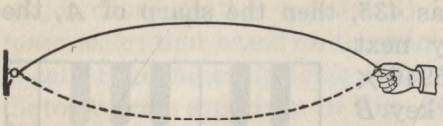


FIG. 486.—String vibrating as a single segment, giving its lowest tone or fundamental

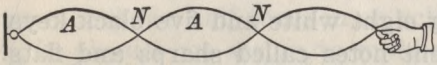


FIG. 487.—String vibrating in three segments, showing nodes and antinodes

518. Fundamental, Overtones, and Harmonics. When a string vibrates as a single segment, it gives off its lowest or fundamental tone. The *fundamental* of any sounding body is the tone of lowest pitch which the body is capable of giving. When a vibrating string breaks up into segments, as shown in Fig. 487, it gives off tones which are higher in pitch than the fundamental; these higher tones are called *overtones*.

In Fig. 488 we have excellent photographs of a string vibrating successively in two, three, and five segments. In the first case (two segments) the pitch of the note emitted is twice as high as the fundamental, that is, one octave above the fundamental; in the second case

(three segments) the note is three times that of the fundamental; and in the third case (five segments)

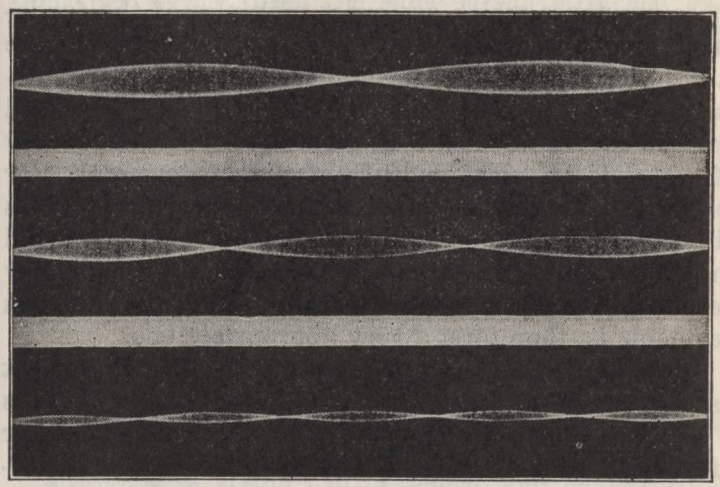


FIG. 488.—Photographs of a string vibrating successively in two, three, and five segments

the pitch is five times that of the lowest note (fundamental).

The illustrations shown in Fig. 488 represent segmented vibrations which are simple exact multiples of

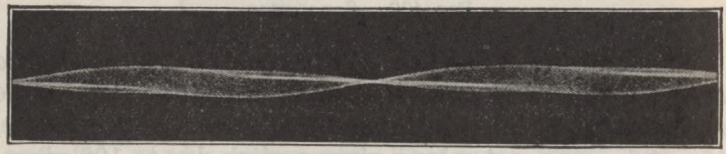


FIG. 489.—Photograph of a string producing a complex tone

the vibration producing the fundamental. A string, however, may, and frequently does, vibrate in a very complicated manner, Fig. 489, producing complex tones. The

overtones resulting from these complex vibrations do not always combine with the fundamental to produce tones which are pleasing to the ear. It has been found that those overtones which are exact multiples of the fundamental, that is, overtones the frequencies of which are 2, 3, 4, 5, times the fundamental combine with the fundamental to produce harmonious tones, and hence are called harmonics. A *harmonic* is an overtone the frequency of which is an exact multiple of the fundamental. For example, if we sound together such a series of notes either with tuning forks or on the piano, it will be found that they combine together so harmoniously that it may be difficult to distinguish one from the other. The term harmonic is, therefore, a very appropriate one.

519. The Sonometer. The *sonometer* is an instrument used in studying the laws of vibration of strings. It is a hollow wooden box upon which is stretched a number of strings. A spring balance or weight is used to determine the stretching force acting on the string, while a small wooden "bridge" is employed to vary its length. By means of an instrument similar to the one shown in Fig. 490, the following laws may be verified.

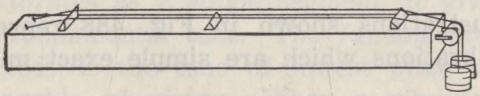


FIG. 490.—Sonometer

520. Laws of Vibration of Strings. I. *The pitch of a string is inversely proportional to its length;* that is, the shorter the string the higher the pitch. For example, if a string one meter long gives a pitch due to 100 vibrations per second, then the same string half a meter in length, the stretching force being constant, will give a pitch an octave higher than the first, that is, a pitch due to 200 vibrations per second.

II. *The pitch of a string is directly proportional to the square root of the stretching force.* If a string stretched by a force of one kilogram makes 100 vibrations per second, then the same string stretched by a force of 4 kilograms will make 200 vibrations per second; that is, $100/200 = \sqrt{1}/\sqrt{4}$.

III. *The pitch of a string is inversely proportional to the square root of the mass per unit of length.* Thus for a given length and stretching force, the lighter the string the higher will be the pitch.

These three laws may be expressed mathematically by the following equation:

$$n = \frac{1}{2l}\sqrt{\frac{F}{m}}$$

in which n is the frequency, l the length of the string in centimeters, F the stretching force in *dynes,* and m the mass of the string in grams *per centimeter of length.*

Example. A string 1 m in length, and having a total mass of 2 grams, is stretched by a force of 2 kg. Find the frequency of the string when it is sounding its fundamental. *Solution*: The length $l = 100$ cm; $F = 2000 \times 980$ dynes; $m = 2/100 = 0.02$ grams per cm. Then the frequency

$$n = \frac{1}{2 \times 100}\sqrt{\frac{2000 \times 980}{0.02}} = 49.5 \text{ vibrations per sec.}$$

Vibration of Air in Pipes

521. Vibration in Pipes. Musical instruments may be divided into two great classes: (a) stringed instruments and (b) wind instruments. The notes which are produced by wind instruments are due to the vibration of air columns within them. We commonly speak of a wind instrument as a pipe.

The various kinds of wind instruments differ chiefly

in the mode of excitation of vibration of the enclosed column of air. The part of the instrument in which the vibrations are excited is called the mouthpiece. Mouthpieces are of three types: (a) Those in which the air is blown across a sharp edge or across an opening, as in the flute or organ pipe. This form of excitation may be illustrated by blowing across the mouth of a small bottle, or better still, a discharged cartridge shell. A common tin whistle illustrates fairly well the principle of construction of the organ pipe. (b) Those in which the air is forced through an opening which is partly closed by an elastic tongue or reed, as in the cabinet organ and the accordion. An excellent illustration of the reed instrument is that furnished by the common "mouth organ," or harmonica. (c) Those in which the air is forced through a slit, as for example that formed by the lips in playing the cornet, or the opening of the vocal cords in the production of voice.

522. Open and Closed Pipes. An *open pipe* is one that is open or free at both ends. In Fig. 491 there is shown an open organ pipe; at one end there is an open mouthpiece, at the other an open or free passage.

In Fig. 492 we have a *closed pipe.* There is at one end an open mouthpiece, similar to that of the open pipe; the other end, however, is closed.

In Fig. 493 there is shown a set of four organ pipes, which when sounding their fundamentals produce a major chord; that is, they sound the notes *do, mi, sol, do'.*

With reference to its fundamental, *an open pipe gives a pitch an octave higher than a closed pipe of the same length.* This may be demonstrated in the following way: Blow gently into the mouthpiece of an open pipe, Fig. 491; the pipe will give forth its lowest or fundamental tone. Now place the hand over the open end of the pipe, thus converting it into a closed pipe; it will now sound

a tone which is one octave lower than that given by the open pipe, provided of course that the fundamental is sounded in each case.

The pitch of a given pipe depends upon its length; in general, the shorter the pipe the higher the pitch. The relation of the length of a pipe to its pitch may be strikingly demonstrated by using a pipe into one end of which is fitted a movable piston, Fig. 494. If the piston be placed at the upper end, and a steady stream of air be blown into the mouthpiece, a tone having a definite pitch will result. If now the plunger be moved downward, the pitch will rise, becoming very shrill as the plunger approaches the mouthpiece. By a single downward and upward motion of the piston, the pitch may be made to rise and fall, illustrating the rise and fall of the pitch of a siren whistle.

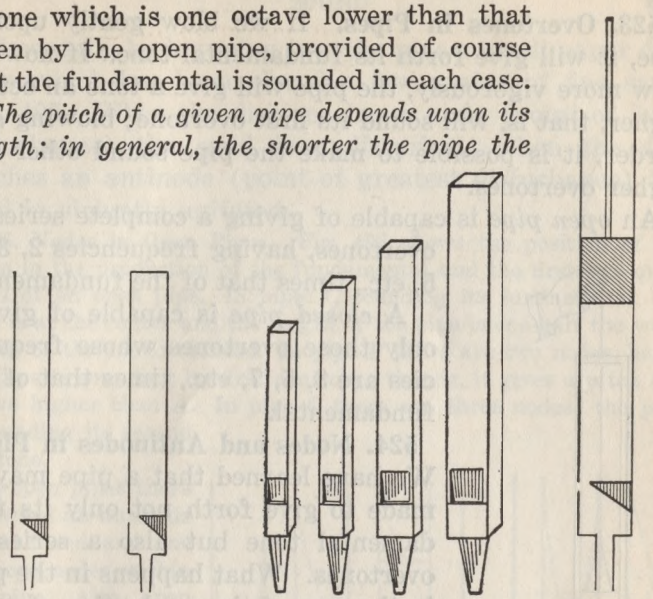


FIG. 491 FIG. 492 FIG. 493 FIG. 494

Open and closed pipes

The length of an open pipe is *one-half* the wave length of the fundamental sound emitted.

The length of a closed pipe is *one-fourth* the wave length of the fundamental sound emitted.

523. Overtones in Pipes. If we blow gently upon a pipe, it will give forth its fundamental tone. If now we blow more vigorously, the pipe will give a tone an octave higher, that is, will sound its first overtone; blowing still harder, it is possible to make the pipe sound other and higher overtones.

An *open pipe* is capable of giving a complete series of overtones, having frequencies 2, 3, 4, 5, etc., times that of the fundamental.

A *closed pipe* is capable of giving only those overtones whose frequencies are 3, 5, 7, etc., times that of the fundamental.

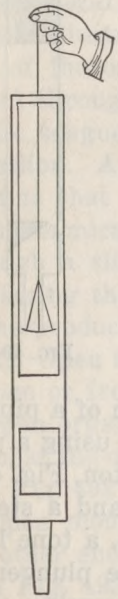


FIG. 495
Experimental determination of the position of a node in a pipe

524. Nodes and Antinodes in Pipes. We have learned that a pipe may be made to give forth not only its fundamental tone but also a series of overtones. What happens in the pipe in the case of the production of overtones is somewhat analogous to that which occurs in a string when it breaks up into segments and gives off its overtones. The air column within the pipe may be made to break up into segments forming nodes and antinodes.

We have said that a node in a pipe is somewhat similar to that in a string. A node in a pipe, however, differs from a node in a string in one very important particular. A node in a string is a point of least motion; *a node in a pipe is a point not only of least motion but also of greatest change of pressure.*

The presence of nodes in pipes may be determined ex-

perimentally by lowering into the pipe a small paper disc upon which has been placed a few grains of fine sand, Fig. 495. When the disc reaches a node (point of least disturbance) the sand is undisturbed; when the disc reaches an antinode (point of greatest disturbance) the sand is violently agitated.

525. Nodes in Open Pipes. Fig. 496 shows the position of the nodes in the production of the fundamental and the first two overtones of an open pipe. In pipe *A*, sounding its fundamental, the node is at the center and the length of the pipe is one-half the wave length of the sound emitted. In pipe *B* there are two nodes, hence the pipe is sounding its first overtone; that is, it gives a pitch one octave higher than *A*. In pipe *C* there are three nodes; the pipe is sounding its second overtone.

In open pipes there is always an antinode at each end and one or more nodes within the pipe. Also, open pipes are capable of sounding in addition to their fundamental tone their first, second, third, fourth, fifth, etc., overtones.

526. Nodes in Closed Pipes. Pipe *D*, Fig. 497, represents a closed pipe sounding its fundamental; *E* represents the same pipe giving its first overtone, the vibration rate of which is three times that of the fundamental; *F* is sounding the next overtone, which is five times the vibration rate of *D*. Closed pipes are capable of giving only the first, third, fifth, etc., overtones.

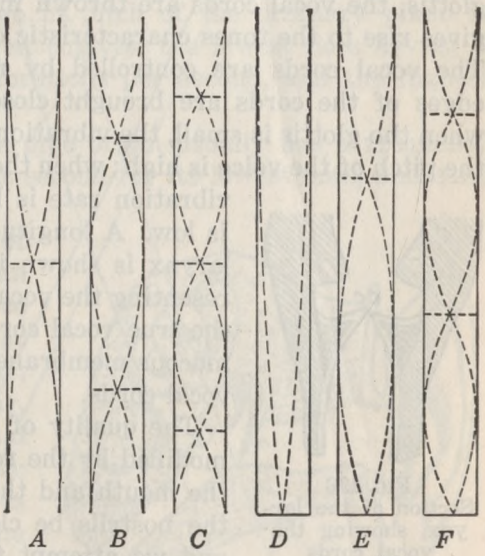


FIG. 496
Nodes and antinodes in open pipes

FIG. 497
Nodes and antinodes in closed pipes

Speech and Hearing

527. The Vocal Organs. In the upper portion of the throat there is a hard projecting mass commonly called "Adam's apple," which constitutes a part of the *larynx*, a box-like chamber formed of cartilages. The larynx is the primary organ of speech and song, and is of all musical instruments the most wonderful, both on account of its simplicity as well as for its extreme delicacy of range. Across the upper end of the larynx there are stretched two muscular membranes, called the *vocal cords*. The narrow opening *cc* between the vocal cords is called the *glottis*. When the air is forced through the glottis, the vocal cords are thrown into vibration, which gives rise to the tones characteristic of the human voice. The vocal cords are controlled by muscles. When the edges of the cords are brought close together, that is, when the glottis is small, the vibration rate is high, hence the pitch of the voice is high; when the glottis is large, the vibration rate is low, hence the tone is low. A longitudinal section of the larynx is shown in Fig. 498, *cc* representing the vocal cords. Just above the true vocal cords are two folds of mucous membrane, *ff*, called the false vocal cords.



Fig. 498
Section of the larynx, showing the vocal cords

The quality of the human voice is modified by the resonance cavities of the mouth and the nasal passage. If the nostrils be closed by the fingers and we attempt to recite the words "The moon is beaming," we will have strikingly illustrated the effects of changing the resonant properties of the nasal cavity.

528. Pitch Limits. *Human voice.* The range of pitch

for the human voice in singing is from 60 vibrations per second for a low bass to about 1300 vibrations per second for a very high soprano. The range for the ordinary individual voice is about two octaves. The range of the voice of the trained singer, however, is frequently somewhat above two octaves. The lowest note which a bass voice can reach is about two octaves below middle *C*, or about 60 vibrations per second. The highest note which the ordinary soprano can reach is about two octaves above middle *C*, or somewhat above 1000 vibrations per second. The pitch of a woman's voice is in general about twice as high as that of a man's voice. It must be noted, however, that the range of each is practically the same, namely, about two octaves.

Piano. The range in pitch of the ordinary piano is somewhat over seven octaves, the lowest note having a frequency of 27 vibrations per second, and the highest 4138.

Organ pipe. The pipe organ usually has a frequency of 16 vibrations per second for its lowest note, and 4138 for the highest.

529. The Organs of Hearing. The human ear consists of three parts or divisions: (a) the external ear, (b) the middle ear, and (c) the internal ear, Fig. 499. The external ear includes the concha and the tube or canal terminating in a thin membrane called the drum membrane or tympanum *T*. The middle ear contains a

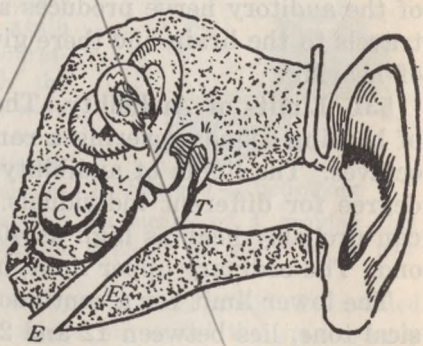


FIG. 499.—Mechanism of the ear

chain of little bones, the function of which is to transmit vibrations from the tympanum to the internal ear. The middle ear communicates with the mouth chamber by means of a tube called the Eustachian tube *E*, the function of which is to equalize the pressure between the air in the middle ear and that on the outside. The internal ear is placed deep in the skull and consists of three parts: (a) the vestibule *V*, (b) the semi-circular canals *S*, three in number, and (c) the cochlea, or snail shell *C*, all of which are filled with a watery fluid. In this watery fluid of the internal ear we find the nerve fibers, which spread out from the auditory nerve.

530. How We Hear. The process of hearing begins with the transmission of sound waves to the drum of the ear. These vibrations, it will be remembered, consist of condensations and rarefactions which produce upon the drum membrane changes of pressure, causing it to vibrate back and forth. The vibrations are then transmitted by the ossicles of the middle ear to the fluids of the internal ear, whence it is transmitted to the fibers of the auditory nerve. This disturbance of the end organs of the auditory nerve produces a nervous impulse which travels to the brain and there gives rise to the sensation of hearing.

531. Limits of Audibility. The human ear is capable of hearing sounds covering a range in pitch of about ten octaves. The limits of audibility vary in quite a marked degree for different individuals. Some exceptional ears can probably hear as high as 40,000 vibrations per second. The ordinary upper limit, however, is about 30,000.

The lower limit for a continuous sound, that is, a musical tone, lies between 12 and 20 vibrations per second, the general consensus of opinion being that the lower limit for a musical tone is 16 vibrations per second. Of course the ear can hear sounds which have a vibration

frequency lower than 16 per second, but they are heard as separated or discontinuous sounds.

Talking Machines

532. Introductory. To the inventive genius of Thomas Edison we owe the two instruments which have done so much to enlarge the field of educative media as affecting the ear and eye, namely, the phonograph and the moving picture. The phonograph, invented by Edison in 1877, originally consisted of a rotating cylinder which recorded the movements of a diaphragm provided with a small needle by means of indentations in a sheet of tin foil supported over a spiral groove on a metal cylinder. In later machines the metal cylinder with its tin foil covering was replaced by a wax cylinder or disc. In the modern dictaphone, Fig. 500, the cylindrical wax receiver is still employed. Machines of the "Dictaphone" and "Ediphone" types are designed to take the place of a stenographer in business offices. The business man dictates into the receiver of the machine, the moving diaphragm with its tracing needle recording the sounds of his voice on the surface of a rotating wax cylinder. Later the typist allows the cylinder to rotate again under the needle, and the vibrating diaphragm gives forth the sounds of the words which were imparted to it. When the record has been received and written out on the typewriter, the surface of the wax cylinder is made smooth by means of a graving tool attached to the machine, and the cylinder is ready for a new record.



Fig. 500.—Dictaphone

A modification of the cylinder type of phonograph was invented by Bell and Tainter; this machine was called the graphophone. Later on Berliner, of Washington, D. C., introduced the method of etching the original record on a zinc disc, producing the gramophone type. In the preparation of the gramophone disc, which is now the familiar disc record, the vibrating needle of the dia-

phragm is allowed to vibrate over the surface of a rotating greased zinc disc. The trace of the needle in the film of grease on the surface of the disc is then etched by means of an acid, and then a copper mold is made of the etched surface of the zinc by means of the electrotyping process. By means of this copper record hundreds of facsimiles of the original record may, by pressure and heat, then be made upon the bakelite disc of commerce.

In accordance with common practice, we shall use the term "phonograph" to designate the various types of speech recording and reproducing machines.

533. The Phonograph. So far as simplicity of construction and perfection of execution are concerned, the phonograph is closely allied to the organs of human speech and hearing. This instrument is provided with a mouthpiece across one end of which there is placed a thin disc, usually of mica, to the center of which there is attached a needle. When it is desired to prepare a record, the diaphragm is adjusted so that the stylus or needle rests lightly upon the rotating cylinder or disc, as the case may be. Sound waves entering the receiver cause the diaphragm to vibrate, and this in turn causes the needle to trace out in the soft face of the rotating receiving surface an irregular path corresponding to the character of the sound waves impressed upon it. As explained in Art. 532, hard rubber copies of the receiving surface are now prepared in quantity, and the records are ready for use. When this hard surfaced record is caused to rotate under the needle or stylus of the machine, the vibrating diaphragm gives forth the same wave forms as were impressed upon it, and talking, singing, or instrumental music results. A photograph of the tracing on a portion of the surface of a record of orchestral music is given in Fig. 501.

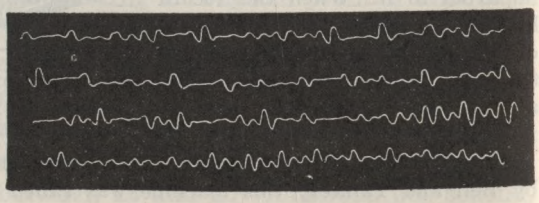


FIG. 501.—Photograph of the tracing of a portion of a record of orchestral music

The quality of tone produced by a phonograph depends upon (a) the condition of the surface of the record, (b) the kind of needle used, (c) the speed of rotating of the disc, and (d) the

resonant properties of the cavities and framework of the instrument.

Analysis of Musical Sounds

534. The Phonodeik. Many instruments have been devised for the optical analysis of sounds. One of the most sensitive instruments of this sort is that known as the *phonodeik*, designed and perfected by Professor D. C. Miller. An outline section of this wonderfully sensitive instrument is shown in Fig. 502, in which h

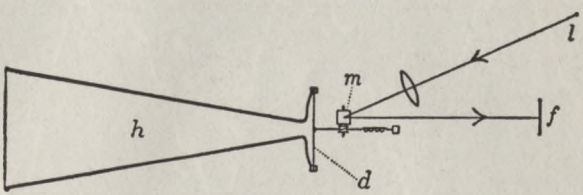


Fig. 502.—Outline section of the phonodeik

is the receiving horn; d is a sensitive diaphragm of thin glass, placed at the end of the horn; behind the diaphragm is a minute steel spindle mounted in jeweled bearings, to which is attached a tiny mirror m. A ray of light l is focused by means of a lens and reflected by the mirror m to a moving photographic film f. When the diaphragm d moves due to the action of a sound wave, the mirror m vibrates, thus causing the reflected beam of light to trace a record of the sound wave on the moving photographic film at f.

The musical instrument which gives the simplest and purest tone is the tuning fork.

535. Tones from Stringed Instruments. The best known stringed musical instruments are the violin, piano, harp, guitar, banjo, and mandolin.

The violin. The strings of the violin are caused to vibrate by the to-and-fro motion of the bow; these vibrations are transmitted through the bridge and body of the instrument to the air. The great advantage of the violin over all the other orchestral instruments in expressiveness is due to the control by means of fingers and bow which the performer has over the tone production. In Fig. 503 there is shown a photograph of a typical violin tone. Variations in bowing of the strings change the regularity of the vibrations and thus produce a continually changing wave form.

The piano. The vibrations in a piano originate in the strings which are struck by felt-covered hammers controlled by the keys.

While the vibrations originate in the strings, most of the sound comes from the sounding board to which the strings are attached. The piano is perhaps one of the most expressive instruments upon which one person can play, and hence it is undoubtedly the most popular musical instrument. Upon this instrument the skillful

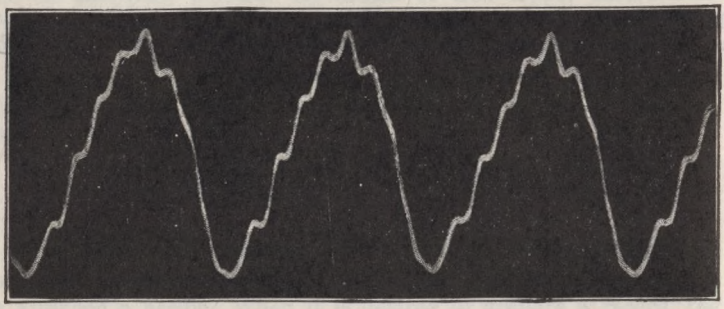


FIG. 503.—Photograph of a typical tone from the violin

player can produce wonderful varieties of tone color in chords and groups of notes, resulting in music which is full, rich, and varied. In relation to tone quality as produced on a piano, it is believed that the so-called "artistic touch" of the musician consists in slight variations in the time of striking the different keys, as well as variations in the strength of the blows, and that tone quality is determined by purely physical and mechanical considerations. "The correctness of this argument," says Prof. Miller, "is further supported by the mechanical piano players, which attempt to reproduce the characteristics of individual pianists. The more highly developed such instruments become, the more nearly they imitate hand playing in musical effects; in many instances the imitation is practically perfect, and I believe that in the near future the automatic piano will reproduce all of the effects of hand playing."

536. Wind Instruments. *Organ pipe.* The organ pipe is one of the simplest of wind instruments. The pipe organ usually has a frequency of 16 vibrations per second for its lowest note and 4138 for its highest. An organ pipe giving a note due to 16 vibrations per second is nominally 32 feet long; that having a frequency of 4138 is 1.5 inches in length. There are a few organs in the world which have pipes 64 feet in length. A pipe of this length gives only 8 vibrations per second; such sounds, however, are hardly to

be classified as musical tones, as they are heard as separated or discontinuous sounds.

The flute. The flute gives the simplest tones of any orchestral instrument. In Fig. 504 we have a photograph of one of a selected

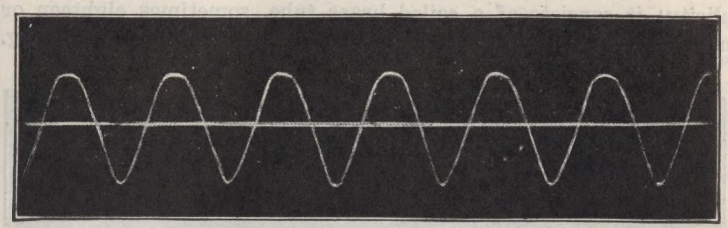


FIG. 504.—Photograph of the simple tone of a flute

series of tones of a flute. Note the symmetrical outline of the particular curve, which in many respects resembles the simple curve given by a tuning fork. The flute tone may be described as sweet and tender. The poet Sidney Lanier, himself an accom-

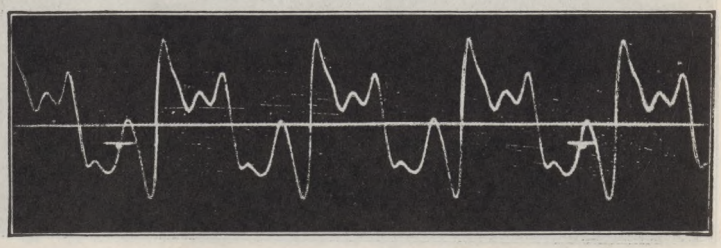


FIG. 505.—Photograph of the tone of a clarinet

plished flutist, in his poem "The Symphony" describes the tone of this musical instrument in the following beautiful lines:

"But presently
A velvet flute note fell down pleasantly
Upon the bosom of that harmony,
And sailed and sailed incessantly,
As if a petal from a wild rose blown
Had fluttered down upon the pool of tone."

The clarinet. The clarinet generates sound by means of the vibration of a single bamboo reed in the mouthpiece. The body of

the instrument has a uniform cylindrical base, at the lower end of which there is a bell shaped opening. Fig. 505 shows a photograph of a characteristic tone of the clarinet. The tones of this instrument possess a beautiful melting musical quality.

The French horn. The horn is an instrument of extreme simplicity; it consists of a coiled brass tube, sometimes eighteen or more feet in length, bearing at the end a large flaring opening.

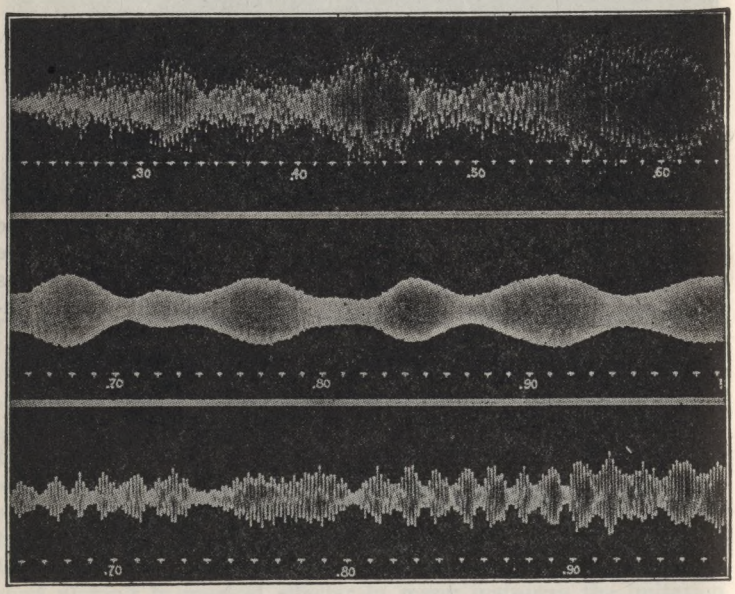


FIG. 506.—Photographs of a reproduction of the sextette from "Lucia di Lammermoor," showing voice curves as produced by Caruso (top), Tetrazinni and Amato

The low sounds of the horn are rich in overtones, containing the largest number yet found in any musical tone.

The human voice. The sound of the voice originates in the vibrations of the vocal chords. The pitch of the voice is controlled largely by the tension of the vocal cords, while the quality depends upon the resonance effects of the vocal cavities. In Fig. 506 we have a section of photographs of the sound waves created by the phonograph record reproducing the sextette from "Lucia di Lam-

mermoor" as sung by six famous operatic stars. The wider parts of the upper picture are due to accented syllables as sung by Caruso. The curve of the second line is from the soprano voice of Tetrazinni, the fluctuations in width being due to a slight tremolo. The third curve is produced by the soprano, and the baritone Amato, singing softly.

SUMMARY

1. Velocity of sound depends upon the elasticity of the medium through which the wave passes and upon the density of the medium. But since a change of temperature may affect the elasticity or the density, or both, we may say that the velocity of sound depends upon three factors, namely, elasticity, density, and temperature.

2. Wave length depends upon two factors, namely, velocity and frequency, as expressed by the equation, $l = v/n$.

3. Pitch depends upon one factor, frequency. The pitch of a sound is that characteristic which is determined by the number of vibrations or waves reaching the ear per second.

4. A musical sound is one which has regularity of pitch and which produces an agreeable effect upon the ear.

5. A musical interval is the ratio of the pitch of one note to that of another. The ratio 1:1 is called unison; the ratio 1:2, an octave.

6. A major triad is made up of three notes having the ratios 4:5:6. A major chord is a major triad together with the octave of the lowest note, that is 4:5:6:8.

7. A minor triad is made up of three notes having the ratios 10:12:16. A minor chord is a minor triad together with the octave of the lowest note, that is 10:12:16:20.

8. A diatonic scale consists of eight notes having definite ratios.

9. In physics the standard of pitch is $C = 256$ vibrations per second. In music the standard is $A = 435$ vibrations per second; this is called the international musical standard of pitch.

10. Sounds may differ in three respects: (a) in loudness, (b) in pitch, and (c) in quality. The quality of a sound is that characteristic which enables us to determine the musical instrument from which it comes. Quality depends upon the form of the sound wave.

11. A node in a vibrating string is the point of least motion; an antinode is the point of greatest motion.

12. When a string vibrates in a single segment, it gives off its lowest tone or fundamental; when it vibrates in more than one

segment, it gives off overtones. The pitch of the overtones is always higher than that of the fundamental tone.

13. A harmonic is an overtone the frequency of which is an exact multiple of the fundamental.

14. Laws of vibration of strings: I. Pitch is inversely proportional to the length of the string.

II. Pitch is directly proportional to the square root of the stretching force.

III. Pitch is inversely proportional to the square root of the mass per unit length.

15. The three laws of vibration of strings may be stated in equational form, thus

$$n = \frac{1}{2l}\sqrt{\frac{F}{m}}$$

16. Variation of pitch in pipes: I. The pitch of the fundamental of an open pipe is an octave higher than the pitch of the fundamental of a closed pipe of the same length.

II. The pitch is inversely proportional to the length of the pipe; that is, the shorter the pipe the higher the pitch.

17. Relation of length of pipe to wave length. (a) The length of a closed pipe is one fourth the wave length of the fundamental.

(b) The length of an open pipe is one half the wave length of the fundamental sound emitted.

18. Overtones in open and closed pipes: (a) An open pipe is capable of producing a complete series of overtones, having frequencies 2, 3, 4, 5, etc., times that of the fundamental.

(b) A closed pipe is capable of producing only those overtones which are odd multiples of the fundamental, that is, 3, 5, 7, etc.

19. A node in a pipe is the point of least motion and greatest change of density. An antinode is a point of greatest change of motion and least change of density. There is always an antinode at the open end of a pipe.

20. Range of pitch of the human voice: The range of the ordinary human voice is about two octaves. The pitch of the lowest note for a bass voice is about 60; that of a high soprano, 1300.

21. Limits of audibility: The ordinary human ear is capable of hearing musical tones between the pitch ranges of from 20 to 30,000.

REVIEW EXERCISES

A

1. Two forks C and C' are sounding at a time when the temperature is 20° C. The frequency of C is 256; the frequency of C' is an octave above C. (a) What is the frequency of C'? (b) What is the velocity of the sounds given off by the two forks?

2. What are the wave lengths of the sounds given off by the forks C and C' (Ex. 1)?

3. Distinguish between loudness and pitch.

4. Explain how a siren may be used to determine the pitch of a sounding body, a church bell, for example.

5. A current of air is blown against the disc of a siren having a row of 30 holes, while the disc is making 120 revolutions per minute. (a) What is the pitch of the resulting tone? (b) If the speed of the siren is doubled, how will the pitch be affected?

6. An A fork makes 426 vibrations per second. (a) What is the frequency of a tuning fork whose pitch is one octave lower? (b) one octave higher?

7. Distinguish between a noise and a musical sound.

8. What is the distinction between a major chord and a minor chord?

9. Name in order the eight notes of the diatonic scale. Are the ratios from note to note of the diatonic scale equal?

10. How many notes per octave are there in the tempered scale of the piano? What do the black keys represent?

11. Make a drawing to illustrate nodes and antinodes in the case of the vibration of strings.

12. What is the distinction between a fundamental tone and an overtone?

13. (a) In what three ways may sounds differ? (b) Define and illustrate what is meant by quality of sound.

14. How will the pitch of a string be affected (a) if its length be doubled? (b) if its tension be quadrupled? (c) if the mass be increased nine times?

15. If we increase the amplitude of vibration of a string, what effect has this upon (a) the amplitude of the wave? (b) the loudness? (c) the pitch?

16. How does the pitch of a pipe vary with its length?

17. A plunger is thrust into a pipe which is sounding a given

note. As the plunger moves into the pipe, what change in pitch occurs? Explain.

18. What is the relation of the pitch of an open pipe to that of a closed pipe of the same length?

19. An open pipe of given length is sounding its fundamental. Suppose that one end is stopped by means of a card. How will the pitch be affected?

20. Explain the condition of the vocal cords when the pitch of the voice is (a) high, (b) low.

21. In what way is the quality of the human voice modified?

22. (a) What is the octave range of hearing? (b) What are the pitch limits of audibility for the ordinary ear?

B

23. When the temperature is 22° C, a man singing gives forth a sound the wave length of which is 7.09 ft. (a) What is the frequency of the note? (b) What note in the diatonic scale is he singing?

24. (a) What is the ratio from one key to the next on the piano? (b) What name is given to a system of tempering in which the ratios are all equal?

25. (a) State the laws of vibration of strings. (b) Explain each term of the equation $n = \frac{1}{2l}\sqrt{\frac{F}{m}}$

26. A given string stretched by a force of 16 lb. makes 256 vibrations per second. What will be the vibration rate if the stretching force be increased to 25 lb.?

27. A string 1 m long makes 512 vibrations per second. What will be the frequency of the string if its length be decreased by 25 cm?

28. Define node in a pipe, and tell wherein it differs from a node in a string.

29. How may the presence of nodes in pipes be determined experimentally?

30. What overtones may be obtained from (a) an open pipe? (b) a closed pipe?

31. (a) What are the pitch limits from low bass to high soprano? (b) What is the octave range of the ordinary singing voice?

C

32. What is the wave length of international *A* when the speed of sound is 1131 ft. per second?

33. The frequency of the lowest note on the piano is 27. Find the wave length for this note when the temperature is 20° C.

34. The frequency of the highest note on the piano is 4138. Find the wave length for this note when the temperature is 68° F.

35. A whistle of a train which at rest makes a certain number of vibrations per second which strike the ear of a person who is standing still. Suppose now that the train moves toward the person. (a) Will the number of vibrations falling upon his ear per second be increased or diminished? (b) How will this affect the pitch? (c) How dó you account for the rise of the pitch of a locomotive whistle which is rapidly approaching?

36. (a) Make a sketch of the ear, and describe how changes of pressure due to condensations and rarefactions of sound waves are transmitted from the external to the internal ear. (d) Describe the Eustachian tube, and explain briefly its function.

NOTE: For additional problems, see Appendix.

CHAPTER XVIII

LIGHT

NATURE OF LIGHT

537. Introductory. The sun is not only the most familiar source of light, but it is also the earth's chief source of energy. All the energy of plant and animal life is directly traceable to the sun, as is also the energy furnished by wind and water power. The energy of our coal fields, too, is nothing more than the stored-up energy of the sun's rays of by-gone ages. Now the sun is some 93,000,000 miles from the earth, and the question may be asked: How does this energy get to the earth across the enormous gulf of intervening space? In general, two theories have been advanced to explain the transmission of the sun's energy; namely, (a) the emission theory, and (b) the wave theory.

The *emission or corpuscular theory,* which was advocated by Sir Isaac Newton, assumed that the energy of the sun is carried through space by tiny particles or corpuscles which are shot off from that body with an enormous velocity, and which carry energy somewhat as a bullet carries energy from a gun.

The *undulatory or wave theory,* which was first definitely stated by Huygens, a Dutch physicist, in 1678, assumes that the energy of the sun is transmitted through space by means of waves, through a medium called the ether, which is supposed to fill all space.

Light is that form of radiant energy which enables us to see things; it is propagated through space by a kind of wave which is caused by the vibration of electrons. Only a few years ago we were in the habit of saying

that Newton's corpuscular theory was entirely wrong. We are not so sure of that today, however. While it is true that the wave theory, as proposed by Huygens, explains most of the phenomena associated with light, yet recent developments arising from the application of the quantum theory of radiation to problems in light indicate the necessity of far reaching modifications in our simple wave theory. A discussion of these modifications, however, at the present time would manifestly be out of place in an elementary text. We shall therefore devote our attention to a consideration of the application of the wave theory to the phenomena of light.

538. **Wave Theory of Light.** The wave theory of light assumes that the energy of the sun and other light giving bodies is transmitted through space by means of waves which travel outward from the source, in a medium of uniform density, in the form of concentric spherical waves. Such a propagation of light waves may be represented on a flat surface by a series of concentric circles as shown in Fig. 507, in which L is the source of light. The line LA is drawn to represent the direction of propagation of a given set of wave fronts, as ab.

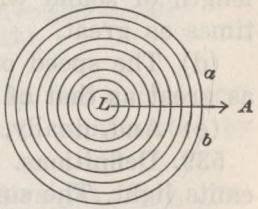


FIG. 507.—Light waves travel outward in all directions from the source L

It is important here to note that only a part of the energy that comes from the sun by means of ether waves is capable of affecting the eye and thus producing the sensation of sight. Ether waves from the sun may be divided into three classes: (a) relatively long waves which produce the sensation of heat; (b) short waves which affect the eye, producing the sensation of sight; and (c) very short waves which do not affect the organs of sight but which are capable of giving rise to chemical

reactions, such for example as those which may affect a photographic plate. Now while it is not always customary to do so, we shall in this text, for the sake of clearness, refer to those waves which affect the eye as light waves.

Light waves differ from sound waves in several very important particulars, as follows:

(a) Light waves are transmitted through a hypothetical medium called ether; sound waves are transmitted only through matter. Light can be transmitted through a vacuum; sound cannot.

(b) Light waves are transverse; sound waves in fluids (air and water) are longitudinal.

(c) Light waves are exceedingly short; sound waves are relatively very long. For example, waves of yellow light are only 0.0000589 centimeter in length, while the length of sound waves in air are more than 2,000,000 times as great.

(d) The speed of light is more than a million times as great as that of sound.

(e) And, finally, light affects the eye; sound, the ear.

539. Definitions. (a) A *luminous* body is one which emits light. The sun and the stars are luminous bodies; so, too, is a candle flame or the glowing filament of an incandescent lamp. Bodies which shine by light other than their own are called *illuminated* bodies. The moon is a good illustration of an illuminated body. A *transparent* body is one which allows light to pass through it readily; that is, it is one which we can see through, as, for example, window glass. A *translucent* body is one which allows only a part of the light to pass through without permitting objects to be distinctly seen, as in the case of "frosted" glass. An *opaque* body is one which does not allow light to pass through it. It must be noted in this connection that the terms transparent, trans-

lucent, and opaque are used in a purely relative sense. Wood, for example, is opaque, yet a shaving from a pine board may be obtained so thin as to be almost transparent. Gold may be beaten out into sheets so thin as to be translucent.

(b) *Optics* is that branch of physics which treats of light. *Light* may be defined as a vibratory motion which is capable of affecting the organs of sight.

(c) A *ray, LA,* Fig. 507, is a line drawn to represent the direction of propagation of a wave front, *ab.* A *beam* of light is a number of parallel rays, Fig. 508. A pencil of light is a number of rays passing through a point called the *focus.* In Fig. 509 there is represented

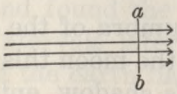


FIG. 508
Beam of light

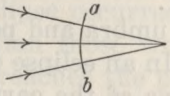


FIG. 509

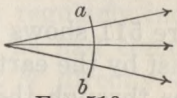


FIG. 510
Converging and diverging pencils of light

a *convergent pencil* of light and in Fig. 510 a *divergent pencil.* The wave front in each case is represented by the curved line *ab.*

540. Shadows. Light travels in straight lines, provided the medium through which it is propagated is *homogeneous,* that is, provided the medium has the same density and elasticity throughout. The fact that light travels in straight lines is seen when a beam is passed into a darkened room in which there are particles of dust floating in the air. The marksman in sighting his gun, the surveyor in adjusting his instruments, and the mechanic in much of his work, all take account of this fundamental property of light, namely, that it travels in straight lines.

If an opaque object be held in front of a luminous body there appears in the rear of the opaque body a dark

space called a *shadow*, which has three dimensions: length, breadth, and thickness. That portion which appears on the wall or screen is called a section of the shadow. A shadow may consist of two parts, the *umbra* or complete shadow and the *penumbra* or partial shadow.

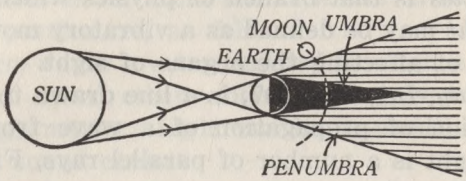


FIG. 511.—Shadow, showing umbra and penumbra

Figure 511 shows the umbra and penumbra of the shadow cast by the earth. In an eclipse of the moon this body passes through the cone of the earth's shadow, entering the penumbra first and leaving it last.

451. The Velocity of Light. For a long time it was thought that light was transmitted through space in-

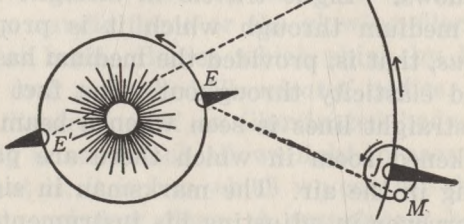


FIG. 512.—Illustrating Roemer's method of determining the speed of light

stantaneously. In 1675, however, Roemer, a young Danish astronomer, who was studying the satellites or moons of Jupiter at the observatory at Paris, came to the con-

clusion that light travels through space with a definite finite speed. According to his determination, the velocity of light (transmission in a given direction) is about 186,000 miles per second. Roemer's famous discovery came about somewhat as follows: The planet Jupiter has several moons which revolve about it as our moon does about the earth. The inner one of these satellites passes into the shadow cast by Jupiter, Fig. 512; that is, it is eclipsed once on an average every 42 hours, 28 minutes, 36 seconds. Roemer calculated in advance the exact time that each eclipse should occur for a number of positions of the earth in its orbit. He took a series of observations as the earth moved from E away from Jupiter and found the eclipse occurred at regularly increasing intervals later than those which he had computed. When the earth had reached the point E' directly across its orbit from Jupiter, the eclipse occurred about 1000 seconds later than the computed time. He concluded that this apparent delay in the time of the eclipse was due to the time required for light to travel across the earth's orbit a distance of about 186,000,000 miles. Now 186,000,000/1000 = 186,000; that is to say, *the velocity of light is* 186,000 *miles per second.*

Roemer's wonderful discovery was received with but little favor by the scientific men of his day; indeed, it was practically disregarded for over fifty years until other methods were discovered for measuring the speed of light and his figures were thus verified.

542. Velocity of Light in Different Media. According to Newton's corpuscular theory, light should travel faster in a dense medium, such as water or glass, than in a rare medium; according to the wave theory, on the other hand, light should have the greater velocity in a rare medium. Therefore, when in 1850 Foucault, a French physicist, measured, by means of a rotating mir-

ror, the speed of light in air and water, and found thereby that its velocity is greater in air than in water, he thus established on an apparently firm basis the undulatory or wave theory which is now generally accepted.

Undoubtedly the most precise determination of the velocity of light was that made by Prof. Michelson of the University of Chicago. Michelson's value gives to light a velocity of 299,860 kilometers per second. Both Foucault and Michelson demonstrated by laboratory methods that the speed of light is greater in a vacuum than in air, and greater in air than in water or glass.

EXERCISE. **1.** Michelson's value of 299,860 kilometers per second is equivalent to how many miles per second?

543. Images Formed through Small Apertures. If a candle or other luminous body be placed in front of a small aperture in an opaque screen, Fig. 513, there will be formed upon a second screen *S* an inverted image of the

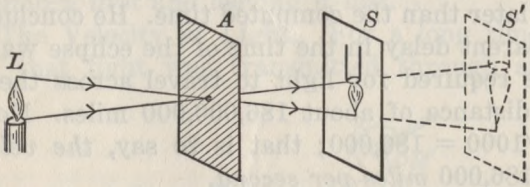


FIG. 513.—Inverted images formed through a small aperture

body. The following points with respect to the formation of images through apertures are to be noted: (a) The image is inverted and reversed; that is, it is upside down with respect to the object and the right side of the image corresponds to the left side of the object. This is due to the fact that rays of light from the different points of the object pictured on the screen cross as they pass through the aperture.

(b) The size of the image depends on the distance of the screen from the aperture.

(c) The smaller the opening the less the illumination

of the image but the greater the distinctness of outline; on the other hand, as the aperture becomes larger the illumination of the image increases but its distinctness diminishes, due to the overlapping of the rays.

The principle of the formation of images through small apertures is made use of in the employment of the so-called *pinhole camera* in photography. This apparatus

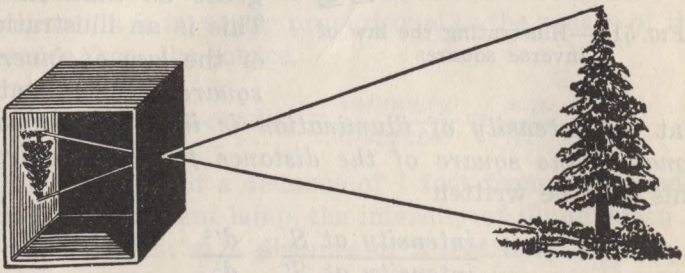


FIG. 514.—Pinhole camera

consists of a small box painted black on the inside to prevent undue reflection of the light which is admitted through a small aperture or "pinhole." The photographic film is placed in the rear of the box in such a way as to receive the image of the object outside, as illustrated in Fig. 514. A camera of this type is very inexpensive and may be employed in photographing landscapes and other scenes when long exposures may be secured.

544. Intensity of Illumination. By *intensity* of illumination we mean the quantity of light per unit area. Experience teaches us that the farther one is away from a given source of light, the less intense the illumination becomes. This is due to the divergence of the rays, as shown in Fig. 515.

Consider screen S to be one foot from the source of light L and screen S' to be two feet from the source. Now the quantity of light which falls upon each screen

is the same. The rays falling on S', however, are spread out over four times as much space as on S; therefore the intensity of illumination on a unit surface of S' is only one-fourth as great as that on S. This is an illustration of the *law of inverse squares*, which states that the *intensity of illumination is inversely proportional to the square of the distance from the source.* This may be written

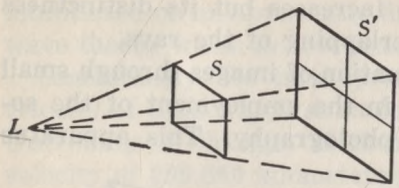


FIG. 515.—Illustrating the law of inverse squares

$$\frac{\textit{intensity at } S}{\textit{intensity at } S'} = \frac{d'^2}{d^2}$$

in which d is the distance of screen S from L, and d' is the distance of S' from L.

EXERCISES. **2.** (a) How does the quantity of light on screen S compare with that on S'? (b) How does the intensity of illumination on S compare with that on S'?

3. (a) Suppose that S is 2 feet from L, and S' 4 ft. How does the intensity of illumination on screen S compare with that on S'? (b) How will the intensities on the two screens compare if S' be placed at a distance of 3 ft. from L?

545. Candle Power and Foot-Candle. The unit of intensity of a source of light is the *candle power;* the unit of intensity of illumination at a given distance from the source is the *foot-candle.*

The *standard of candle power* is the light emitted by a sperm candle 7/8 inch in diameter and burning at the rate of 120 grains per hour. While the sperm candle is the ultimate standard of candle power, it is seldom used any more in actual practice. A much more convenient

and practical standard is a tested incandescent lamp which has been carefully standardized in terms of the standard candle. Our electrical standards of candle power are kept at the Bureau of Standards, Washington, D. C.

A *foot-candle* is the intensity of illumination at a distance of one foot from a light of one candle power. Since according to the law of inverse squares intensity of illumination is inversely proportional to the square of the distance from the source,

$$\textit{foot-candles} = \frac{\textit{intensity}}{\textit{distance}^2} = \frac{\textit{c.p.}}{\textit{ft.}^2}$$

For example, at a distance of 1 foot from a 32 candle power incandescent lamp, the intensity of illumination is 32 foot-candles; at a distance of 2 feet it is $32/2^2 = 8$

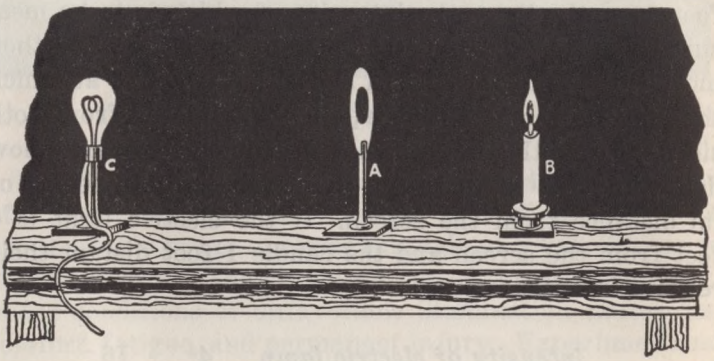


FIG. 516.—Bunsen photometer

foot-candles; and at a distance of 4 feet, $32/4^2 = 2$ foot-candles.

546. The Photometer. A *photometer* is an instrument for measuring the intensity of a source of light in terms of candle power.

The Bunsen photometer, Fig. 516, consists in principle of a paper screen *A*, upon which there is a small circular oil or grease spot. If the paper with the grease spot upon it be held up to the light, in front of an open window, for example, so that the paper is between the light and the eye, the grease spot will appear lighter in color than the paper because it is more transparent. Now if the paper disc be held up in front of a dark body, such as a blackboard, the grease spot will appear darker than the paper, because it reflects less of the light than does the paper. When the intensity of illumination on both sides of the paper disc is equal, the difference in color between the two faces of the grease spot is reduced to a minimum.

To determine the intensity of light by means of this apparatus we proceed somewhat as follows: A standard candle is placed at a given distance from the screen, 1 foot, say; the light, the intensity of which is to be measured, is placed on the other side of the screen and then moved back and forth until a position is found at which the spot on the screen is equally illuminated from both sides. From the law of inverse squares we may now determine the intensity of the given light in terms of its distance from the screen. If the electric light is 4 feet from the screen and the candle 1 foot distant their relative intensities are

$$\frac{\textit{intensity of electric lamp}}{\textit{intensity of candle}} = \frac{4^2}{1^2} = \frac{16}{1}$$

This means that the intensity of the lamp is 16 candle power.

547. The Foot-Candle Meter. In measuring intensity of illumination use is made of the *foot-candle meter*. This

instrument is provided with a screen upon which there is a row of translucent spots illuminated from below by a small electric lamp, the intensity of which may be kept at a constant value by means of a voltmeter and rheostat adjustment. The manner of using the meter is as follows: (a) Adjust the rheostat until the voltmeter indicates that the meter lamp is getting the required voltage; (b) expose the screen to the light the intensity of which is to be measured; (c) select on the screen the spot that appears to be of the same brightness as the screen surface; (d) read on the scale the number of foot-candles which corresponds with this spot. The intensity of illumination as recorded by the meter scale of Fig. 517 is 10 foot candles.

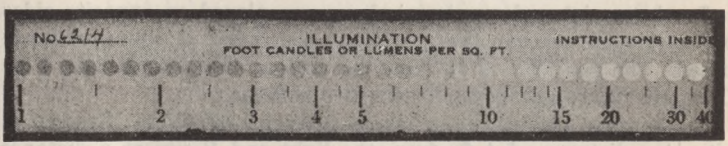


FIG. 517.—Foot-candle meter

548. Suitable Illumination. In strong light the pupil of the eye contracts and in weak light it expands. We are thus able to use our eyes throughout a varying range of illumination. The continued use of the eyes, however, under conditions of unfavorable illumination causes discomfort, fatigue, and permanent injury. Experiment and experience show that eye comfort and efficiency demand for each kind of eye work a certain minimum illumination. Also, since illumination costs money it is desirable for reasons of economy to regulate the intensity according to the purpose for which illumination is provided. The following illumination values were taken from tables recently compiled.

Suitable Illuminating Values

Living rooms	1 to 2 foot-candles
Auditoriums, Shops, rough work	2 to 5 "
Class rooms, Libraries, Offices, Shops, fine work	5 to 10 "
Engraving, Drafting, Sewing, dark goods	10 to 20 "

EXERCISES. **4.** A lamp placed 3 feet from a screen gives an illumination equal to that of a standard candle placed at a distance of 1 ft. from the screen, lamp and candle being on opposite sides. Find the candle power of the lamp.

5. A gas jet and an electric lamp are placed on opposite sides of the screen so that the spot upon the latter is equally illuminated. The gas jet is 2 ft. from the screen; the electric light 6 ft. Compare the candle power of the two sources.

6. A given electric light has an intensity of 32 c.p. Student *A* sits at a distance of 4 ft. from the light and *B* sits at a distance of 8 ft. Find the intensities of illumination in foot-candles for these two positions.

REFLECTION AND REFRACTION OF LIGHT

549. Reflection. If a beam of light from a source *A*, Fig. 518, falls upon a polished surface, it will be reflected to *C*. The line *BD* is drawn perpendicular to the surface of the mirror and is called a *normal*. The angle *ABD*, formed by the incident ray and the normal, is called the *angle of incidence;* the angle *D B C*,

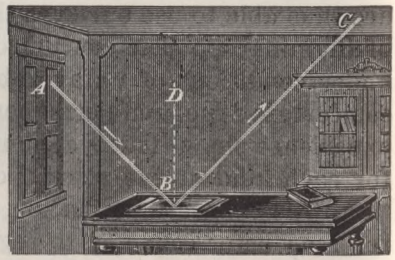


FIG. 518.—The angle of reflection is equal to the angle of incidence

formed by the reflected ray and the normal, is the *angle of reflection.* The angle of incidence is usually represented by the letter *i,* the angle of reflection by the letter *r.*

Law of reflection: *The angle of reflection is equal to the angle of incidence, and both lie in the same plane.*

550. Regular and Diffused Reflection. If a beam of light falls upon a smooth surface, it will be *regularly reflected,* as shown in Fig. 519. If this light be received by the eye, the latter will see not the reflecting surface, but the source of light. For example, light reflected from a mirror, from the surface of highly polished furniture, or the surface of still water shows this property.

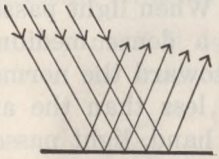


FIG. 519

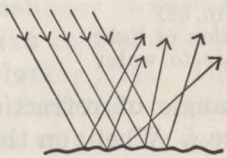


FIG. 520

Regular and diffused reflections

If, on the other hand, a beam of light falls upon a body having an irregular surface, such as a piece of white paper or the wall of a room, the light is scattered in all directions, Fig. 520. This is called *diffused light.* It must not be imagined, however, that diffused light does not obey the laws of reflection, for it does. The angle of reflection for each ray is in every case equal to the angle of incidence. The scattering of the rays is due to the irregularities in the reflecting surface. It is by means of this diffused light that we are enabled to see clearly the outline of bodies. If every object possessed a polished surface, we would see only the light of the source and would not be able to make out clearly and

definitely the outline and nature of the reflecting surface.

551. Refraction. If a beam of light from A in air be allowed to fall upon the surface of a transparent medium, such as water, Fig. 521, a part will be reflected in accordance with the laws of reflection, and a part will be refracted to A'. *Refraction* is the bending of a ray of light due to passing from a medium of one density to that of another density. The ray AO is the *incident ray,* and OA' the *refracted ray.* The angle i is the *angle of incidence,* r the *angle of refraction,* and d the *angle of deviation.* When light passes from a rare to a dense medium, it is refracted toward the normal; that is, the angle of refraction r is less than the angle of incidence i. When, on the other hand, light passes from a dense medium (water or glass) to a rare medium (air), Fig. 522, it is refracted away from the normal; that is, r is greater than i.

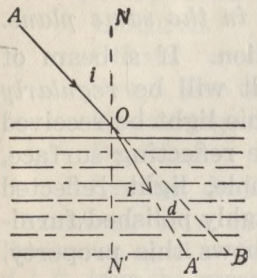


FIG. 521
Refraction of light from air to water

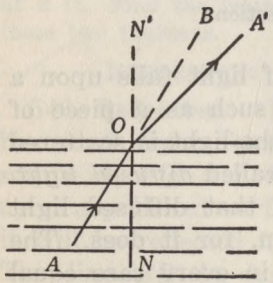


FIG. 522
Refraction of light from water to air

Refraction may be explained as follows: Imagine a series of wave fronts advancing in the direction AO, Fig. 523. When the portion of the wave marked b strikes the surface of the water, its speed is retarded; a, therefore, moves faster than b, hence the wave front is bent downward, as shown, and the direction of the motion is changed from AO to OB. After a wave once passes into a medium both portions of the

wave front, a and b, move forward with the same speed; therefore the line of direction OB is a straight line. Refraction of light is due to a change in the speed of one portion of a wave front as it passes from one medium to another medium of different density.

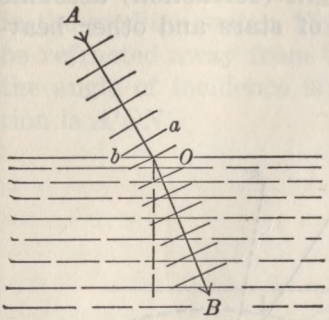


FIG. 523
Explanation of refraction

552. **Examples of Refraction.** Place a coin on the bottom of an empty cup so that it lies just inside the range of vision as bounded by the side of the vessel. Now pour water carefully into the cup until it is full; the coin comes gradually into view as the surface of the water rises. This is due to the fact that rays of light from the coin to the eye are bent (refracted) at the surface of the water, as shown in Fig. 524. The eye in looking along the line appears to see the coin in an elevated position. In like manner when we look vertically downward at the bottom of the vessel filled with water, the apparent distance from the surface to the bottom is less than the real depth; indeed the bottom appears to be in a position about three-fourths of the real depth.



FIG. 524
Refraction of light

Again if a stick be placed in a slanting position in water, it appears to be bent at the surface, due to the bending of the rays of light in their passage from the relatively dense water to the rare air. This apparent bending at the surface may be beautifully demonstrated

by observing a spoon placed in a tumbler which is partly filled with water, Fig. 525.

The bending of the rays of light (refraction) accounts for the apparent displacement of stars and other heav-



FIG. 525
An illustration of refraction

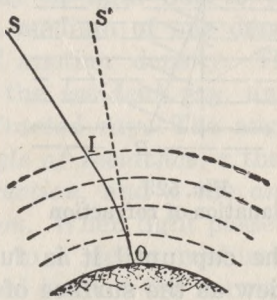


FIG. 526
Refraction due to the passage of light through layers of air of different densities

enly bodies, Fig. 526. The ray of light from the star S is bent (refracted) as it passes through successive layers of the atmosphere so that an observer on the surface of the earth at O sees the star as if it were at S'. The apparent position of the star is higher than the real position. It thus frequently happens that the sun is visible while yet actually below the horizon.

The mirage is a phenomenon frequently observed in deserts, in which the traveler sees the image of distant objects, such as palm trees, etc., usually by refraction. The explanation of the mirage lies mainly in the fact of the refraction of light through layers of air of different density. A phenomenon somewhat similar to the mirage of the desert is occasionally seen at sea in still, hot

weather, in which the image of a distant ship appears in the sky.

553. Critical Angle. Consider a source of light at A in the medium water, Fig. 527. A ray from A to O will be refracted away from the normal to A'. For this ray, the angle of incidence is AON and the angle of refraction is $A'ON'$.

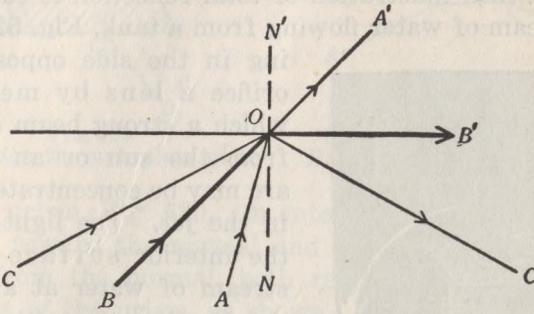


FIG. 527.—Critical angle, BON

Suppose now that the source of light be moved farther to the left of the normal NN' to such a position that the incident ray BO will just graze the surface OB'. In this case the angle of refraction $B'ON'$ is 90°, and the angle BON is called the critical angle. The *critical angle* is an angle of incidence such that the angle of refraction is 90°. When we speak of a critical angle we always think of the light as passing from a relatively dense to a less dense medium, as from water or glass to air.

If the source of light be moved still farther to the left of the normal, as to C, for example, the incident ray CO will be reflected to C'. That is to say, a diver at C' looking at the surface at O would see the bottom at C.

When the angle of incidence of a ray of light passing from a dense to a rare medium is *greater* than the critical angle, we have *total internal reflection*. The smaller

the critical angle of a given substance, the greater is the total internal reflection. The diamond, for example, has a small critical angle, and therefore, most of the light falling upon the surface from the interior suffers total internal reflection. The large proportion of light which is totally reflected in the diamond gives rise to the great brilliancy of this gem.

A beautiful illustration of total reflection is furnished by a stream of water flowing from a tank, Fig. 528, having in the side opposite the orifice a lens by means of which a strong beam of light from the sun or an electric arc may be concentrated within the jet. The light strikes the interior surface of the stream of water at an angle greater than the critical angle and thus suffers total internal reflection. The light is held within the stream until, after repeated reflection, it strikes the bottom of the receptacle, where it shows a bright spot. A goblet held in the stream is filled with bright light, which gives rise to the name "fountain of fire." The brilliant display seen in electric fountains depends upon this principle.

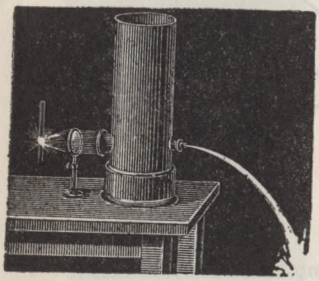


FIG. 528.—The "fountain of fire"

554. Passage of Light through Different Media. *Plate glass.* Let a ray of light from A be incident upon the surface of a piece of plate glass at the point B, Fig. 529. At B the ray is bent toward the normal as it passes into the denser medium, and at C is bent away from the normal as it passes out. The refraction at C is equal to that at B but in the opposite direction. The ray CD is therefore parallel to AB.

When a ray passes through a plate having parallel

faces, its direction is unchanged; it suffers only lateral displacement.

Prism. Let a ray of light from an object at A fall

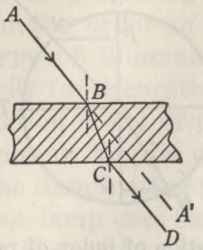


FIG. 529
Refraction through plate glass

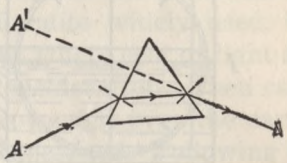


FIG. 530
Refraction through a prism

upon a prism, Fig. 530. On entering the prism the light is bent toward the normal and on emerging, it is bent away from the normal, both refractions being toward the base of the prism, as shown. An eye will appear to see the object at A', as if it were elevated in line with the final refracted ray.

555. Index of Refraction. *The index of refraction* of a medium is the ratio of the velocity of light in air to the velocity of light in the given substance. For example, the ratio of the velocity of light in air is to that in water as 4 to 3; that is, light travels 4/3 as fast in air as in water. The index of refraction of water is therefore $4/3 = 1.33$.

There are several methods of determining the index of refraction for any given substance. One of the simplest (though not the most accurate) is by means of the refraction tank, Fig. 531. A refraction tank is a rectangular glass vessel which is half filled with water, and one face of which is provided with a graduated circle. Over the top of the tank there is placed an opaque screen, having a slit S for the admisson of a beam of

light. A beam of light L, is reflected from a mirror M into the tank at N. This beam of light, MN, falls upon

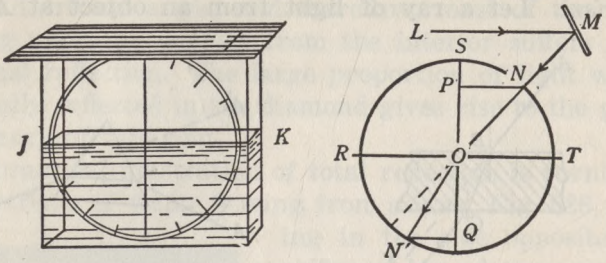


FIG. 531.—Refraction tank for demonstration of index of refraction

the surface of the water at O, and is refracted to N'. Draw the lines NP and $N'Q$ perpendicular to the vertical line SQ. Now it may be shown mathematically that the velocity of light in the air is to that in the water as line PN is to line QN', both of which may be measured. It may be shown experimentally that for *water with respect to air*

$$\textit{index of refraction} = \frac{PN}{QN'} = \frac{4}{3} = 1.33$$

The indices of refraction of a number of substances are given in Table XIX, Appendix. With respect to air the *index of refraction* of

Water is	1.33	Flint glass	1.62
Crown glass	1.52	Diamond	2.47

ARTIFICIAL LIGHTING

556. Introductory. Primitive man used torches, and in some cases crude lanterns holding swarms of fireflies, to guide him when he went forth at night. From that day to the present, one of the problems which man has attempted to solve is the production of an illuminant in which a large part of the energy supplied appears as light

and a small part as heat. Thus far his attempts have been relatively feeble as compared with what nature has accomplished in the case of the firefly, an insect which is capable of producing what is practically a cold light. Next in the order of artificial lights came the candle. This type of illuminant is still quite widely used, but ordinarily for decorative purposes where cost of light and illumination is not of primary consideration. Then came the kerosene lamp, a great improvement over the candle so far as illuminating power is concerned. Following the kerosene lamp we have the gas light, and finally the modern high power electric lamp.

The cost of lighting by any of the modern methods depends to a certain extent upon local conditions. In some localities natural gas is available and is very cheap, and in other places electric energy can be obtained at a very low cost. The cost in any case depends upon the candle power of the lamps used and upon the time they burn. In order then to make comparisons between kinds of lamps, it is necessary to consider a definite amount of lighting, which may be obtained by multiplying the candle power of the lamps used by the number of hours they burn. For example, 1000 candle-hours of lighting may be obtained by burning a 10-candle lamp 100 hours or a 50-candle lamp 20 hours. If the lamps are of the same kind, the cost will be about the same. Calculations of the cost of producing 1000 candle-hours by different lamps are sometimes used in choosing between lamps, but of course it does not necessarily follow that the lamp for which the cost is lowest is most desirable for household use.

557. Distribution of Light. The proper distribution of light with reference to the most satisfactory illumination depends in a large measure upon the kind of shade or reflector used. There are three methods of distribution

which are used with lighting systems. These are (a) the direct method, (b) the indirect method, and (c) the semi-direct method.

The *direct method* is one in which the light shines directly from the source upon the area illuminated. This method gives the maximum amount of light for a given expenditure of energy; it is in general objectionable, however, because of the glare which comes from the lamp or from polished surfaces upon which the light falls.

The *indirect method* is one in which the light is placed within bowl-like reflectors near the ceiling. The light is thrown upward upon the ceiling and is thence reflected as diffused light throughout the room. The indirect system of lighting gives less illumination per unit of light at the source than that furnished by the direct system; it, however, avoids the glare of the direct light, and in addition it gives a soft, pleasing lighting effect throughout the room. A ceiling reflector for the production of indirect diffused illumination is shown in Fig. 532.

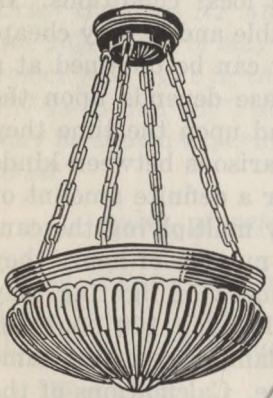


FIG. 532
Reflecting bowl for indirect lighting

The *semi-direct method* is designed to combine the advantages of both the direct and the indirect systems by reflecting part of the light directly upon a limited area and part to the ceiling, from which it is reflected diffusely. In this method of illumination that portion of the light which is intended to fall upon a given area, as upon a table or desk, is frequently transmitted through a translucent screen or shade.

A diagram of the distribution of light from a tungsten electric lamp provided with opal glass reflector is shown in Fig. 533. The curved dotted lines on each side of the lamp represent the distribution of the light when the reflecting shade is removed; the heavy lines below the lamp represent the distribution due to the reflector.

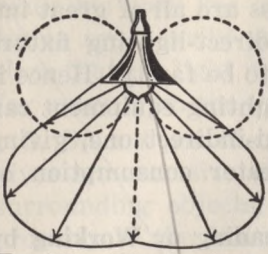


FIG. 533.—Distribution of light by a reflecting shade

With regard to the relative efficiency of direct, indirect, and semi-indirect systems of lighting there has been much argument. There can be no doubt that every time light is reflected or is transmitted through a diffusing medium part of the light is lost. In a properly planned direct-lighting equipment most of the light is reflected only once or not at all before reaching the place where it is used. In indirect lighting the light undergoes at least one additional reflection (at the ceiling), causing a loss which can hardly be less than 25 per cent, and it may be much more. The diffusing shades used in so-called semi-direct lighting throw a large part of the light to the ceiling, where it suffers a similar loss, while that part of the light which goes through the bowl is reduced considerably by absorption. Consequently, if we measure the efficiency of lighting by the fraction of the light which is finally thrown down upon the objects in the room, well-planned direct lighting will always be found to be most efficient. This advantage is accentuated if it is desired to light a particular area, for direct lighting fixtures give most exact control of the distribution of the light, the other methods unavoidably giving a rather widely scattered light.

However, the quantity of light falling upon a given

area is not the only thing which determines whether or not one can see well. The proper diffusion of the light, the illumination of the surroundings, and the protection of the eye from excessive brightness are all of great importance, and in these respects a direct-lighting fixture is more likely than the other types to be faulty. Hence it often happens that a poor direct lighting equipment can be replaced by an indirect, or semi-indirect one, giving a better illumination with no greater consumption of power.

558. **Rules to Be Observed in Reading or Working by Artificial Light.** 1. *Do not work with a flickering light.* In such a light the eye is constantly being called upon to readjust itself and is therefore never properly adapted for the light it receives at any particular instant. Severe fatigue is the result.

2. *Do not use unshaded lamps.* The bare flame, mantle, or filament is injuriously bright, and without shades one's glance is certain to fall upon the dazzling lamp more or less frequently. Moreover, diffused light, that is, light from a considerable surface, usually gives more effective illumination than that from a small source.

3. *Do not judge of illumination by looking at the lamp.* The lamp is not meant chiefly to be seen but rather to enable you to see other things. The lamp which itself looks most brilliant may be for that very reason least effective. The light that comes directly from the lamp to the eye does no good and may interfere very seriously with the useful light which has gone from the lamp to surrounding objects and thence to the eye.

4. *Do not face the light.* The old rule that light should come from above and over the left shoulder is good. With the lamp so placed there is little likelihood of getting the brilliant reflection from table tops, smooth paper, or other shiny material which is so detrimental to good

seeing; there is also little chance of letting the eyes fall upon the lamp itself. The common arrangements of lamps on library tables where the reader faces the lamp, while his book lies on the table, is very bad, unless the lamps are heavily shaded and the book lies at such a slope that the light regularly reflected from the page falls far away from the reader's eyes.

5. *Do not put the lamp too close to your work.* This gives an excessive contrast between the work and the surrounding objects, and the eye does not adapt itself properly to the intensity of light on the work. In factories it has been found that if lamps are removed from individual machines and hung high enough to give a good illumination over the machines and their surroundings, much better work can be done, and the total amount of light needed is often no greater. Similarly, tests have shown that when a desk lamp is used in a room otherwise unlighted the average person needs much more light to read on the desk than he does when the light is furnished by lamps which at the same time illuminate the whole room moderately.

SUMMARY

1. Light differs from sound in the following respects: (a) Sound affects the organs of hearing; light, the organs of sight. (b) Sound is transmitted through ponderable (weighable) matter; light, through a hypothetical medium called the ether. (c) Sound waves in air are longitudinal, consisting of condensations and rarefractions; light waves are transverse. (d) Sound travels in air at the rate of about one-fifth of a mile (1090 ft. at 0° C) per second; the velocity of light is 186,000 miles per second.

2. Images formed through small apertures: (a) The image is real and inverted. (b) The linear size of the image is directly proportional to the distance of the screen from the aperture. (c) The smaller the aperture the less the illumination, but the sharper the image outline.

3. Law of inverse squares: The intensity of illumination at a

given point is inversely proportional to the square of the distance from the source of light.

4. The unit of intensity of a source of light is the candle power. One candle power is an intensity equivalent to the light emitted by a standard candle. The international standard unit of candle power is the light emitted by a sperm candle ⅞ inch in diameter and burning at the rate of 120 grains per hour. For practical purposes incandescent lamps which have been tested in terms of the standard candle are used.

5. The unit of intensity of illumination is the foot-candle. A foot-candle is that intensity of illumination which is equivalent to the intensity at a distance of 1 foot from a standard candle; foot-candles = c.p./ft.^{2}.

6. The angle of incidence is the angle made by the incident ray with the normal; the angle of reflection is the angle made by the reflected ray and the normal.

7. Law of reflection: The angle of reflection is equal to the angle of incidence.

8. Refraction is the bending of a ray of light out of its course as it passes from a medium of a given density to a medium of different density. The angle of refraction is the angle made by the refracted ray and the normal.

9. When light passes from a rare to a dense medium, the refracted ray is bent toward the normal; when, on the other hand, light passes from a dense to a rare medium, the refracted ray is bent away from the normal.

10. The critical angle is that angle of incidence in the denser medium such that the angle of refraction in the rare medium is 90°.

11. The index of refraction of a medium is the ratio of the velocity of light in air to the velocity of light in the given substance.

REVIEW EXERCISES

A

1. Name two theories of the transmission of light, and give the names of the physicists who advanced them.

2. Wherein do light waves differ from sound waves?

3. Define and illustrate (a) luminous body, (b) transparent body, (c) opaque body.

4. Show by means of a sketch the different parts of a shadow. Explain the distinction between umbra and penumbra.

5. Compare (a) the velocity of light with that of sound; (b) the velocity of light in a vacuum with that in air; (c) velocity in air with that in water.

6. Explain by means of a sketch why images formed through small apertures are inverted.

7. Explain what is meant by intensity of illumination, and state and illustrate the law of inverse squares.

8. Two screens are placed at a distance of 2 ft. and 3 ft. respectively from a given source of light. What must be the relative areas of the two screens in order that they intercept the same amount of light?

9. Distinguish between candle power and foot-candle.

10. A 32 candle power light is placed 4 ft. above a writing desk. (a) What is the intensity of illumination on the desk in foot-candles? (b) What would be the illumination in foot-candles if the light were 1 ft. above the desk?

11. Explain how a photometer may be used to measure candle power.

12. What is the distinction between a photometer and a foot-candle meter?

13. Make a drawing to illustrate (a) angle of incidence, (b) angle of reflection. How do these two angles compare?

14. Make a drawing to illustrate (a) regular reflection, (b) diffused reflection.

15. Is a perfectly transparent body visible? Explain.

16. Is a perfectly reflecting surface visible? Explain.

17. Define and explain refraction. Explain how light is refracted with reference to the normal when it passes (a) from a rare to dense medium; (b) from a dense to a rare medium.

18. By means of a diagram explain (a) angle of incidence, (b) angle of refraction, (c) angle of deviation.

19. When is the angle of refraction (a) less than the angle of incidence? (b) greater?

20. A man standing on the bank of a stream wishes to spear a fish which lies in the water below him. Should he strike above or below the apparent position of the fish?

B

21. Explain Roemer's method of determining the velocity of light. What is the velocity of light?

22. (a) In the case of images formed through small apertures, how is the intensity of illumination affected by moving the source of light toward the aperture? (b) How is the intensity of illumination affected by increasing the size of the aperture? (c) What effect does this have on the distinctness of the image?

23. Two sources of light, 8 and 32 c.p. respectively, are placed 6 ft. apart. At what point between them will the illumination produced by them be equal?

24. A book is sufficiently illuminated when held 2 ft. from an incandescent lamp of 8 c.p. How far from a 32 c.p. lamp should the book be placed in order to receive the same illumination?

25. It is desired to find the candle power of a given arc light. It is found that when a standard candle is placed 1 ft. from the screen of the photometer, the latter has to be removed to a distance of 30 ft. from the arc light. Find the candle power of the arc.

26. Make a drawing to illustrate critical angle.

27. How do you account for the brilliancy of the diamond?

28. Trace a ray of light (a) through a plate glass having parallel surfaces; (b) through a prism.

29. Define index of refraction. What does it mean when we say that the index of refraction of water is 4/3? What is the velocity of light in water?

C

30. It is said to take fifty-four and one-half years for light to travel from the North star to the earth. How far away is the North star?

31. The star Arcturus is 600,000,000,000,000 miles from the earth. Suppose that this star were suddenly extinguished today, on what date would astronomers know that this had happened?

32. In Art. 552 it was stated that when one looks vertically downward at the bottom of a vessel filled with water, the apparent depth is ¾ of the real depth. Explain.

33. Considering the index of refraction of crown glass to be 1.5, what is the velocity of light in crown glass?

CHAPTER XIX

LIGHT (Continued)

MIRRORS

559. Mirrors and Images. A *mirror* is a polished surface capable of reflecting light. With reference to their shape, mirrors are of two general classes, plane and spherical. Glass mirrors are made by coating one side of plate glass with an amalgam of mercury. Mirrors of this sort, therefore, possess two reflecting surfaces, the front glass surface and the rear quicksilver surface. Most of the reflected light comes from the rear surface.

Images formed in mirrors are of two kinds, virtual images and real images. Whenever one looks into a plane mirror, a virtual image is observed. The image appears to be as far back of the mirror as the object is in front of it. A *virtual image* is one which is due to the apparent focusing of the rays of light; it appears to be where it really is not. A virtual image appears to be behind the mirror only so long as there is an eye in front of the mirror to receive the reflected rays. When the observer withdraws, the virtual image disappears. Images formed in plane mirrors are always virtual.

A real image is one that is formed by the actual focusing of the rays of light; in other words, a real image is one that actually exists, whether there be an eye present to observe it or not. The formation of a real image may be demonstrated by means of a spherical mirror, as will be shown later.

560. Multiple Images. If an object such as a candle

flame or a gas flame be held close to a plate glass mirror and viewed at an angle, a series of images may be observed, Fig. 534, the second being the brightest of the series. These are called *multiple images,* and their formation is explained by the fact that the light from the object is reflected from the mirror to the eye from both front and rear surfaces of the mirror. The first image which appears is due to the reflection from the surface of the glass. The second, and brightest image, is due to the reflection from the silvered covering on the rear surface of the mirror. The other and fainter images are caused by successive reflections from the two surfaces of the mirror.

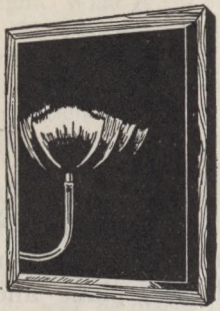


FIG. 534
Multiple images

It is because of the formation of multiple images that glass mirrors cannot be used successfully in certain kinds of scientific work, such as astronomical observations. For this purpose there are employed mirrors which are made of highly polished metal, and which, therefore, have but a single reflecting surface.

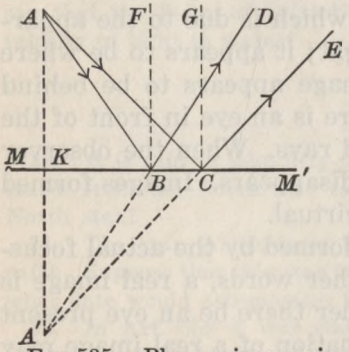


FIG. 535.—Plane mirror; virtual image at A'

561. Images Formed by Plane Mirrors. Images formed by plane mirrors are virtual. A virtual image is one formed by the apparent focusing of the rays of light from an object, Fig. 535. Suppose we have

given an object in front of a plane mirror to find the position and character of the image. Let A be an object in front of the mirror MM'. Draw from the point A two lines incident upon the mirror at B and C. Erect the normals BF and CG. Now draw the lines BD and CE in such a manner as to make the respective angles of reflection equal to their corresponding angles of incidence; that is, make angle $FBD = ABF$ and $GCE = ACG$. Now project the lines DB and EC until they meet, as at A'. An eye placed at DE so as to receive the rays BD and CE will see the virtual image at A'.

If a person stands in front of a plane mirror and observes his image, the following facts may be noted: *First, the image is virtual and appears to be as far back of the mirror as the observer is in front of it; second, it is erect; and third, it is reversed,* that is, as the image faces the observer, the right arm appears on the left side and the left arm on the right side, and so also for all other corresponding parts.

562. Spherical Mirrors. A *spherical mirror* is formed from a portion of a spherical shell. A section of a spherical mirror is shown in Fig. 536. The opening MM' is called the *aperture* of the mirror. A point midway between M and M' is the *vertex* V. The *center of curvature* C of the mirror is the center of the sphere of which the mirror is a part.

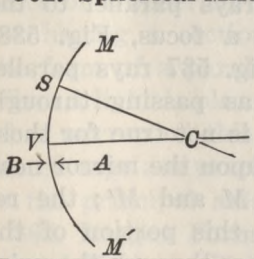


FIG. 536.—Spherical mirror, defining terms:
Center of curvature, C
Vertex, V
Principal axis, CV
Secondary axis, CS
Concave side, A
Convex side, B

A *principal axis* is a straight line passing through the center of curvature and the vertex. A *secondary axis* is any other straight line passing through the center of curvature and cutting the mirror at any point other than at V.

When we consider a spherical mirror from the concave side, it is called a *concave mirror*. When we look at it from the convex side, it is called a *convex mirror*.

The *principal focus* of spherical mirrors is the point at which rays parallel to the principal axis come to a focus, Fig. 537. The principal focus F lies at a point midway between V and C. For *concave* mirrors this focus is *real* and a point, as may be shown by allowing the sun's rays, which are nearly parallel, to fall upon the mirror.

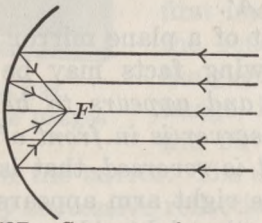


FIG. 537.—Concave mirror

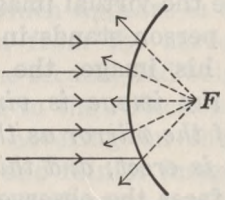


FIG. 538.—Convex mirror

The principal focus of a *convex* mirror is a virtual focus and is the point at which rays parallel to the principal axis appear to come to a focus, Fig. 538.

563. Spherical Aberration. In Fig. 537 rays parallel to the principal axis are shown as passing through the principal focus. This, however, is not true for those rays which fall upon the mirror near the extremities M and M'; the reflected rays for this portion of the mirror do not pass through the principal focus. This failure of part of the light to pass through the principal focus is called *spherical aberration*, Fig. 539.

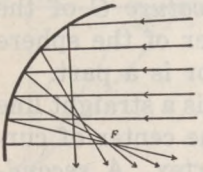


FIG. 539
Spherical aberration

Spherical aberration in the case of a circular reflecting surface gives rise to a curved line of light called the *caustic*. A caustic

curve can be demonstrated experimentally by allowing rays of light to fall upon the concave surface of a strip of polished metal (bright tin) bent in the form of a circular arc, Fig. 540, the reflected light being received on a piece of white paper upon which the strip of metal rests. This same effect also may be seen by allowing the sunlight to fall upon the inside of a gold finger ring placed on a piece of white paper.

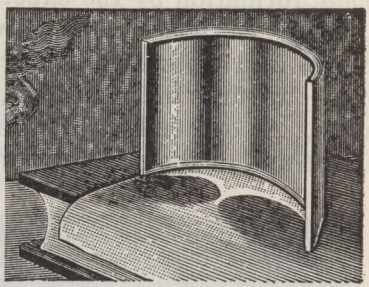


FIG. 540.—Caustic due to spherical aberration of light

In headlights and searchlights where a beam of light having nearly parallel sides is required, it is necessary to use in place of a spherical mirror a mirror having a sectional outline in the form of a parabola, Fig. 541. When a light is placed at the focus of a mirror of this sort, the reflected rays are practically parallel to the principal axis.

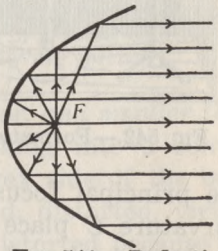


FIG. 541.—Reflection from a headlight mirror

564. Images Formed by Spherical Mirrors. Experimental Method. The character, position, and size of an image formed by a spherical mirror depend upon the position of the object with respect to the mirror. By the *character* of the image, we mean whether it is real or virtual, erect or inverted. The character, position, and size of the image of a given object may be determined in three ways, as follows: (a) by actual experiment; (b) by the graphical method, that is, by making a sketch in

accordance with given conditions, and (c) by means of mathematical equations. We shall now proceed to study image formation by the experimental method.

Case 1, *concave mirror.* Given an object at a distance from a concave spherical mirror greater than the focal length, to find the character, position, and size of the image. At one end of a table place a concave spherical mirror and at the other end a screen, Fig. 542. Between

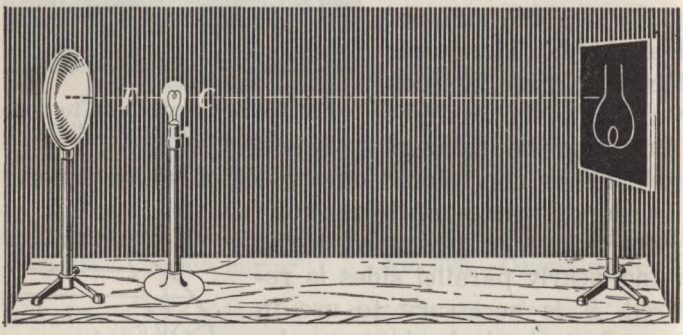


FIG. 542.—Formation of a real image by a spherical mirror

the principal focus F of the mirror and its center of curvature C place an electric lamp or a candle. Move the lamp back and forth until a clear and distinct image is formed on the screen. *The image is real, inverted, and larger than the object.*

If now the positions of the lamp and screen be exchanged, the screen being drawn somewhat forward toward the front of the table and the lamp pushed back so as to allow the light to pass directly to the mirror, there will be formed on the screen a second image which in this case is real, inverted, but smaller than the object. Thus it appears that there are two positions of the lamp and screen which give real images. When the object (lamp) is between the center of curvature C and the

principal focus F, Fig. 542, the image is larger than the object; when the object is beyond the center of curvature, the image is smaller than the object and lies between F and C.

Case 2, *concave mirror.* Given an object in front of a concave spherical mirror at a point between the principal focus F and the mirror, to find the character, position, and size of the image. Hold a concave mirror close to the face so that the eye is between F and the mirror. *A virtual, enlarged image will be seen,* as shown in Fig. 543.

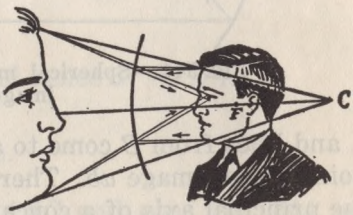


FIG. 543.—Virtual magnified image as seen in a small concave shaving mirror

Case 3, *convex mirror.* Given an object in front of a convex spherical mirror, to find the character, position, and size of the image. *The image is virtual, on the opposite side of the mirror, and smaller than the object,* as may be seen by looking into a convex mirror. When one looks at the convex side of the bowl of a polished silver spoon, a minified, distorted, virtual image may be seen. The image is distorted because the bowl of the spoon is not altogether spherical.

565. Images Formed by Spherical Mirrors. Graphical Method. *Case* 1, *concave mirror.* Given an object AB, Fig. 544, in front of a concave mirror at a distance greater than the center of curvature C, to find the character, position, and size of the image. Draw from A and B to the mirror two straight lines parallel to the principal axis. Since these rays are parallel to the principal axis, they, upon being reflected, pass through the principal focus F. Now from A and B draw lines through the center of curvature C. These rays fall upon the mirror

normally (that is, forming right angles to the surface) and hence are reflected directly back upon themselves. Thus it appears that all lines from *A* come to a focus at

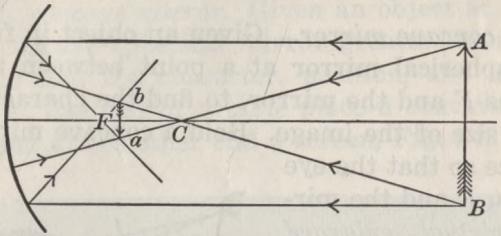


FIG. 544.—Spherical mirror, object *AB* and image *ab*

a, and lines from *B* come to a focus at *b*, forming at this point a real image *ab*. Therefore, when the object is on the principal axis of a concave mirror, at a point beyond the curvature *C*, *the image is real, inverted, smaller than the object, and it lies between the principal focus F and the center of curvature C.* If now we consider that *ab* is the object, located between *F* and *C*, the image will be at *AB*, real, inverted, with respect to *ab*, and larger than the object. That is, when the object is at *AB* the image is at *ab;* and when the object is at *ab*, the image is at *AB*.

In deriving the mathematical equations for the spherical mirrors, we let *d* equal the distance of the object from the vertex *V*, measured on the principal axis of the mirror; *d'*, the distance of the image from the vertex; *r*, the radius of curvature (*VC*) of the mirror; and *f*, the focal length. The focal length $f = \frac{1}{2} r$; that is, $r = 2 f$. For Case 1, Fig. 544, the following equation holds:

$$\frac{1}{d} + \frac{1}{d'} = \frac{2}{r} = \frac{1}{f}$$

in which *d* is the object distance, *d'* the image distance,

r the radius of curvature, and f the focal length. This equation applies when the object is at AB and also when the object is between F and C, that is at ab.

Example. An object AB, Fig. 544, 6 in. in length is placed 5 ft. from the vertex of a concave mirror, the radius of which is 30 in. Find (a) the distance of the image ab from the mirror, (b) the focal length of the mirror, and (c) the size of the image. *Solution*: Reducing all values to inches, we have $d = 60$ in. and $r = 30$ in. Then

(a) $\frac{1}{60} + \frac{1}{d'} = \frac{2}{30}$, and hence $d' = 20$ in.

(b) $f = \frac{r}{2} = \frac{30}{2} = 15$ in.

(c) Since the triangles ABC and abc, Fig. 544, are similar, the size of the image ab is to that of the object AB as their respective distances from the center of curvature C. But the distance of ab from $C = r - d'$, and that of $AB = d - r$. That is

$$\frac{ab}{AB} = \frac{r - d'}{d - r}, \text{ from which}$$

$$\frac{ab}{6} = \frac{30 - 20}{60 - 30} = \frac{1}{3}. \text{ Hence } ab = 2 \text{ in.}$$

Case 2, concave mirror. Given an object on the principal axis of a concave mirror at a point between the principal focus and the vertex V, to find the position and character of the image. Let AB, Fig. 545, be the object. Draw from the points A and B to the mirror lines parallel to the principal axis; these rays are reflected back through the principal focus F. Next draw from A and B lines normal to the surface of the mirror; these rays are reflected back through the center of curvature C. It will be observed that the respective rays from A and B are

divergent, hence they will never come to a focus on the right side of the mirror. *The image is virtual, erect, larger than the object, and on the opposite side of the mirror.*

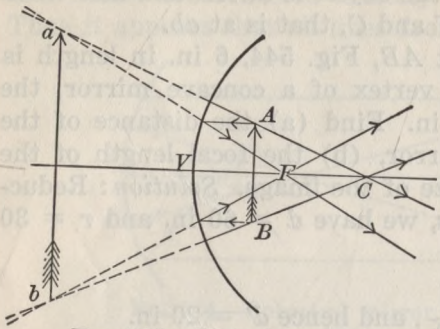


FIG. 545.—Virtual image formed by concave mirror

Since the image is on the opposite side of the mirror from the object, d' is negative and our equation becomes

$$\frac{1}{d} - \frac{1}{d'} = \frac{2}{r} = \frac{1}{f}$$

Example. An object AB is placed on the concave side of a spherical mirror, the focal length of which is 15 in. The image ab is virtual and is situated 30 inches from the convex side. Find the position of the object. *Solution*: Here $d' = 30$ in. and $f = 15$ in., and

$$\frac{1}{d} - \frac{1}{30} = \frac{1}{15}. \text{ Hence } d = 10 \text{ in.}$$

Case 3, *convex mirror.* Given an object AB on the principal axis of a convex mirror, Fig. 546, to find the position and character of the image. Drawing lines from A to B as before, that is, one set parallel to the principal axis and the other set as passing through the center of curvature C, we find that the reflected rays are divergent. Projecting these lines on the opposite side of the mirror gives the location of the virtual image at ab. *In this case the image is virtual, erect, smaller than the object, and located on the opposite side of the mirror.*

Here d', r, and f are all on the opposite side from the object and are therefore all negative. The equation in this case is

$$\frac{1}{d} - \frac{1}{d'} = -\frac{2}{r} = -\frac{1}{f}$$

Example. An object AB, 7 in. in length, is placed 14 in. in front of a convex mirror of focal length 1 ft. Find the size of the virtual image. *Solution*: In order to find the size of ab, we must first find d'. Therefore

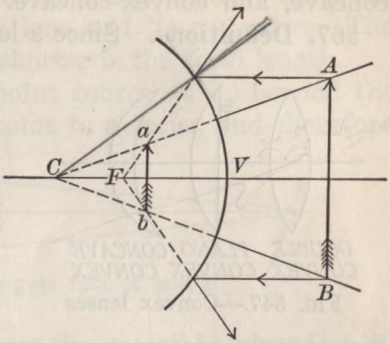


FIG. 546.—Virtual image formed by convex mirror

(a) $\frac{1}{14} - \frac{1}{d'} = -\frac{1}{12}$. Hence $d' = 6.46$ in. Then

(b) $\frac{7}{ab} = \frac{14}{6.46}$, and therefore $ab = 3.23$ in.

LENSES

566. Kinds of Lenses. A *lens* is a transparent medium bounded in general by two spherical surfaces, or one spherical and one plane surface. Lenses are of two general classes, convex and concave.

A *convex lens* is one that is thicker in the middle than at the edge. Convex lenses are divided into three subclasses, as shown in Fig. 547, double-convex, plano-convex, and concave-convex. It is important to note that each one of these sub-classes conforms to the requirement of the convex lens; namely, *it is thicker in the middle than at the edge.*

A *concave lens* is one which is thinner at the middle than at the edge. Concave lenses are likewise divided

into three sub-classes, Fig. 548, double-concave, plano-concave, and convex-concave.

567. Definitions. Since a lens is made up of the inter-

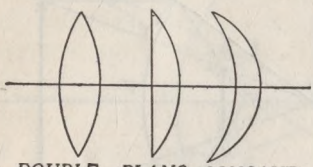


FIG. 547.—Convex lenses

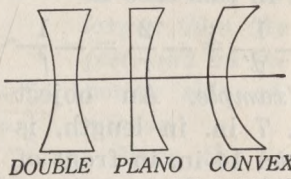


FIG. 548.—Concave lenses

section of two spheres, or of one sphere and a plane, we may speak of the *center of curvature* of the lens as the center of the sphere of which it is a part. Since a lens always has two faces, it therefore always has two centers of curvature.

A straight line drawn through the centers of curvature of a lens is called the *principal axis,* Fig. 549.

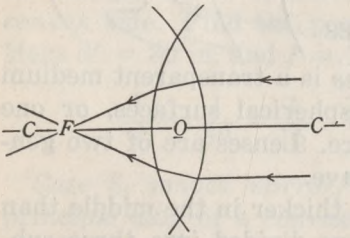


FIG. 549.—Lens, defining terms
Optical center, O
Centers of curvature, CC
Principal axis, COC
Principal focus, F
Focal length, $OF = f$

The *optical center* of the lens O is a point through which a ray may pass without having its direction changed. In double-convex and double-concave lenses, having surfaces of equal curvatures, the optical center is at the center of the lens.

The *principal focus* of a lens is the point on the principal axis at which rays parallel to the principal axis come to a focus. The *focal length* of a lens is the distance measured on the principal axis from the optical center O to the principal

focus F; the focal length of the lens shown in Fig. 549 is $OF = f$. The focal length of a lens depends upon the index of refraction of the lens and its curvature. The greater the curvature the shorter is the focal length.

Rays of light from a point source at A, beyond the principal focus, Fig. 550 come to a focus, and therefore

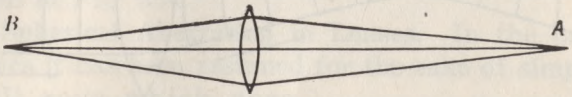


FIG. 550.—Conjugate foci, A and B

form an image at B. If now the object be placed at B, the image will appear at A. Two points, A and B, so situated on opposite sides of a lens that an object at one will form an image at the other are called *conjugate foci.*

568. The Effect of Lenses on Parallel Rays. The effect of a lens in changing the direction of a ray of light is exactly similar to that of a prism; indeed, every section of a lens may be considered as a prism. Rays of

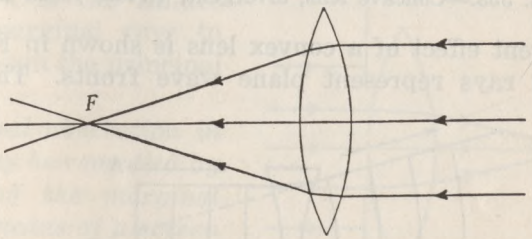


FIG. 551.—Convex lens, convergent, real focus at F

light parallel to the principal axis on entering the lens are therefore bent toward the normal and on leaving it, away from the normal. The general effect of lenses may be stated as follows: (a) *Convex lenses are convergent*;

that is, they cause the rays of light to come to a focus F on the principal axis, Fig. 551. The focus of a convex

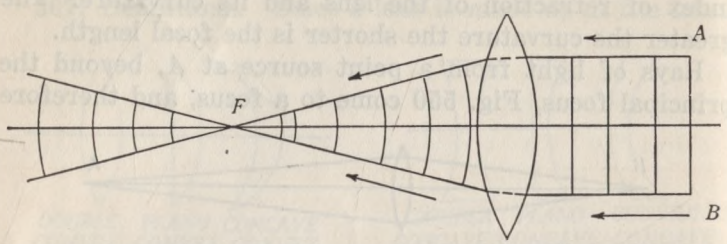


FIG. 552.—Effect of a convex lens on plane wave fronts

lens is real and lies on the opposite side of the lens from which the light comes. Another way of representing the

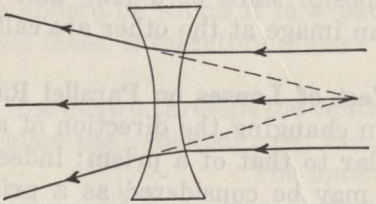


FIG. 553.—Concave lens, divergent, virtual focus at F

convergent effect of a convex lens is shown in Fig. 552. Parallel rays represent plane wave fronts. The plane

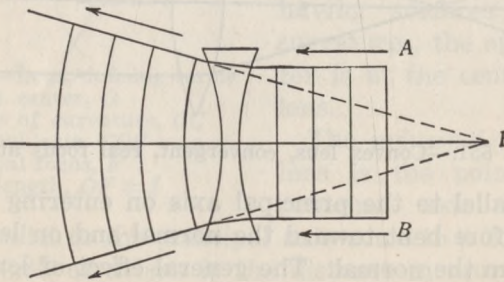


FIG. 554.—Effect of a concave lens on plane wave fronts

wave fronts AB are bent into circular form by the lens, the center of the circles being at the focus F.

(b) *Concave lenses are divergent*; that is, they cause parallel rays to scatter, Fig. 553. The focus of a concave lens is virtual and lies on the side from which the light comes. The effect of a concave lens on a plane wave front is shown in Fig. 554.

569. Spherical Aberration in Lenses. In the preceding topics it has been assumed for the sake of simplicity that all rays which pass through the lens also pass through the principal focus. This, however, is not always the case. In thick convex lenses those rays which pass through near the edge of the lens cross the principal axis nearer the lens than do those rays which pass through near the principal axis, Fig. 555. This failure of all the rays to pass through a common point produces a blurring of the image. *Spherical aberration* is the failure of the marginal rays to pass through the principal focus.

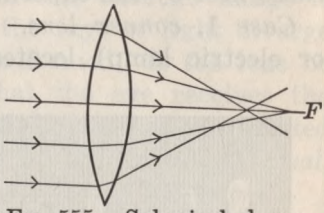


FIG. 555.—Spherical aberration in a lens

Spherical aberration in a lens may be remedied by cutting off the marginal rays by means of a screen or diaphragm, Fig. 556. The use of a diaphragm makes the image sharper in outline, but less bright.

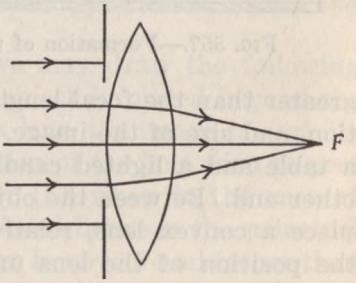


FIG. 556.—Use of a diaphragm for correction of spherical aberration

In large lenses, such as are used in telescopes, spherical aberration is diminished by making the curvature of the

lens less toward the edge, thus tending to bring all parallel rays to the same focus.

570. Images Formed by Lenses. Experimental Method. In relation to the formation of images by lenses, we have, as with mirrors, three general cases to deal with, namely, (a) convex lens, object *beyond* the principal focus F of the lens, (b) convex lens, object *between* the principal focus F and the lens, and (c) concave lens. We shall study the formation of images for these three cases first by the experimental method.

Case 1, *convex lens.* Given an object (candle flame or electric lamp) located at a distance from the lens

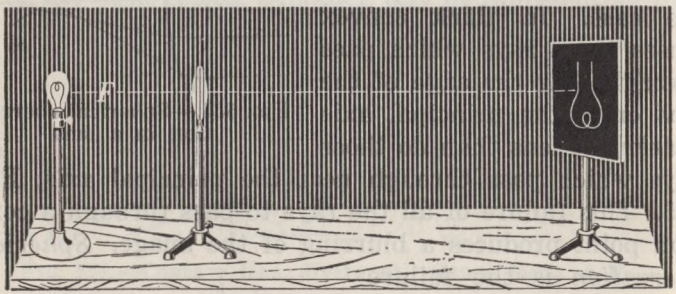


FIG. 557.—Formation of real image by a convex lens

greater than the focal length, to find the character, position, and size of the image. Place a screen at one end of a table and a lighted candle or an electric lamp at the other end. Between the object (candle or electric lamp) place a convex lens, relatively near the object. Adjust the position of the lens until a clear image is focused upon the screen, Fig. 557. *The image is real, inverted, and larger than the object.*

If now we change the position of the lens, placing it as near to the screen as it was formerly near to the object, an image will again be formed on the screen.

The image in this case is real, inverted, but *smaller* than the object. Thus, as in the case of the mirror, the size of the image formed by the lens depends upon its relative position with respect to the screen. When the lens is in its position of focus near the object the image is larger than the object, and when the lens is in its conjugate position of focus near the screen, the image is smaller than the object.

Case 2, *convex lens.* Given an object between the principal focus F and the lens. In this case no image is formed on the screen because the rays of light diverge and therefore no real image is formed. If the lens be used as a reading glass so that the eye receives the divergent rays, an erect magnified image of the printed page may be seen. *The image of the print is virtual, erect, and larger than the actual size of the letters.*

Case 3, *concave lens.* Given an object on the principal axis of a concave lens. No image is formed on the screen, because the rays diverge on passing through the lens. If this lens is used as a reading glass, a virtual, erect, minified image of the printed page may be seen. *In the case of a concave lens, the image is virtual, erect, and smaller than the object.*

From these experiments we may draw the following conclusions: (a) When the object is placed beyond the principal focus F of a convex lens, the image is *real,* its size depending upon the position of the lens with respect to the object and the screen; (b) when in the case of a convex lens the object is placed between the principal focus F and the lens, the image is *virtual* and *larger* than the object; (c) the image formed by a concave lens is always *virtual* and *smaller* than the object.

571. Images Formed by Lenses. Graphic Method. Let us now consider the determination of the character, position, and size of the image by the graphic method, to-

gether with the appropriate lens formula that goes with each case.

Case 1, *convex lens.* Given an object *AB*, Fig. 558, beyond the principal focus *F*, to find the character, posi-

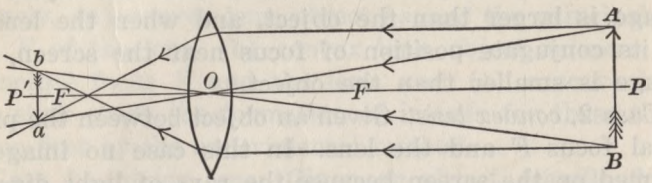


FIG. 558.—Real image formed by convex lens

tion, and size of the image. We select, with reference to each point, two rays the direction of which can readily be determined. With reference to A, these are respectively a ray parallel to the principal axis, which after refraction passes through the principal focus *F* in the direction *Fa*, and the ray *Aa* passing through the optical center. In a like manner we draw from the point *B* two similar rays, one passing through the principal focus and the other passing through the optical center. Thus we have formed at *ab* a graphic representation of the image. *The image is real, inverted, and is located beyond the principal focus F, on the side of the lens opposite to that of the object AB.* The size of the image *ab* depends upon the position of the object *AB*.

If the object is placed at *ab*, the image will be at *AB*.

When *ab* and *F* lie on the opposite side of the lens from the object *AB*, as in Case 1, the lens formula is

$$\frac{1}{d} + \frac{1}{d'} = \frac{1}{f}$$

in which *d* is the object distance, *d'* the image distance, and *f* the focal length of the lens.

Example. An object 4 in. in length is placed 3 ft. from

a convex lens, the focal length of which is 1 ft. Find the position of the image. *Solution*: In this case $d =$ 36 in. and $f = 12$ in. Then

$$\frac{1}{36} + \frac{1}{d'} = \frac{1}{12}. \text{ Hence } d' = 18 \text{ in.}$$

Case 2, convex lens. Given an object AB, Fig. 559, between the principal focus F and the lens, to find the

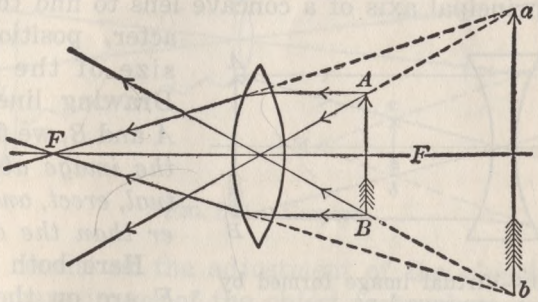


FIG. 559.—Virtual image formed by convex lens

character, position, and size of the image. The rays from AB on the left side of the lens are divergent, hence no real image is formed. If, however, an eye be in a position on the left side of the lens to receive the divergent rays, a magnified virtual image will be seen at ab. *The image is virtual, erect, and larger than the object.* This is the case of the simple microscope (Art. 595).

In this case the image ab has changed positions, being now on the same side of the lens as the object AB. The sign of $1/d'$ is therefore minus, and the lens formula for Case 2 is therefore

$$\frac{1}{d} - \frac{1}{d'} = \frac{1}{f}$$

Example. When an object AB is placed 1 ft. from a

convex lens, the image *ab* appears on the same side as the object and at a distance 30 in. from the lens. Find the focal length of the lens. *Solution*: Here $d = 12$ in. and $d' = 30$ in. Therefore

$$\frac{1}{12} - \frac{1}{30} = \frac{1}{f}. \text{ Hence } f = 20 \text{ in.}$$

Case 3, *concave lens.* Given an object, *AB*, Fig. 560, on the principal axis of a concave lens to find the character, position, and size of the image. Drawing lines from *A* and *B*, we find that *the image ab is virtual, erect, and smaller than the object.*

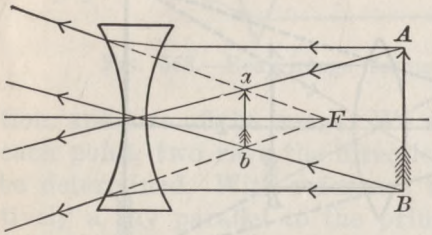


FIG. 560.—Virtual image formed by concave lens

Here both *ab* and *F* are on the object side of the lens, and hence both $1/d'$ and $1/f$ have minus signs. Consequently for Case 3 the lens equation becomes

$$\frac{1}{d} - \frac{1}{d'} = -\frac{1}{f}$$

Example. An object *AB*, 6 in. in length, is placed 3 ft. from a concave lens, the focal length of which is 1 ft. Find the position of the virtual image. *Solution*: Here $d = 36$ in. and $f = 12$ in. Then

$$\frac{1}{36} - \frac{1}{d'} = -\frac{1}{12}. \text{ Hence } d' = 9 \text{ in.}$$

OPTICAL INSTRUMENTS

572. The Camera. The photographer's camera is an instrument used for the printing of pictures by the action of light upon chemically prepared plates or films. It

consists of a light-proof box having an adjustable lens, Fig. 561. The essential features of the operation of the

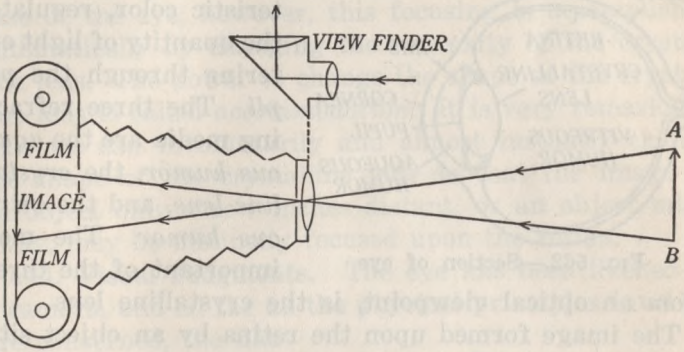


FIG. 561.—Camera

camera consist of the adjustment of the objective lens until a sharp image of the object to be photographed is formed upon the film. The image formed and the object must be at conjugate the foci of the lens. The lens is provided with a diaphragm and shutter to regulate the amount of light admitted and the time of exposure. With a small opening or "stop," greater sharpness or definition in the picture is obtained, but a greater length of time of exposure must be allowed.

The image formed in the camera is real and inverted.

573. The Eye. The human eye is a small camera in which we have a remarkable application of the principle of the formation of real images by a convex lens. The characteristic parts of the eye are shown in Fig. 562. The outer portion of the eye-ball consists of a tough covering called the *sclerotic coat;* below this is the *choroid coat,* containing a black pigment, the function of which is to prevent internal reflection of the light; the *retina* is formed by an expansion of the optic nerve; covering the

front of the eye is the *cornea,* a hard transparent medium; the *iris,* which gives the eye its characteristic color, regulates the quantity of light entering through the *pupil.* The three refracting media are the *aqueous humor,* the *crystalline lens,* and the *vitreous humor.* The most important of the three, from an optical viewpoint, is the crystalline lens.

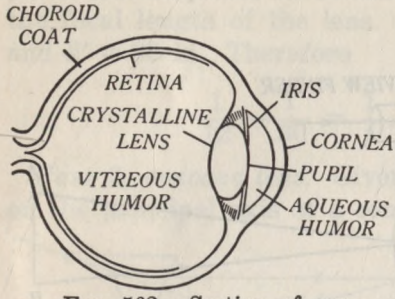


FIG. 562.—Section of eye

The image formed upon the retina by an object situated in front of the eye is real, smaller than the object, and inverted. If a candle be held in front of an eye taken from a freshly killed animal, it is possible under satisfactory conditions to see the inverted image formed upon the retina, as shown in Fig. 563. The question

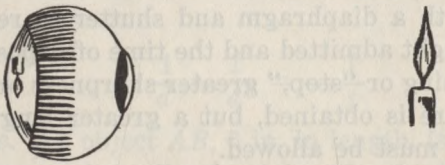


FIG. 563.—Image formed on retina, as seen in the eye of a recently killed ox

naturally arises as to why it is that we can see objects erect when the image upon the retina is inverted. The explanation probably lies in the fact that our judgments of the true position of bodies with respect to each other is based primarily upon experiences other than those of sight. It is said that the blind upon recovering sight often find it necessary at first to use the sense of touch to determine the true position of bodies.

In using the camera it is necessary to focus the instrument by moving either the screen or the lens back and forth so as to bring the image upon the plate. In the case of the eye, however, this focusing is accomplished automatically by changing the convexity of the crystalline lens. The power to change the shape of the crystalline lens is called *accommodation.* It is very remarkable that we can involuntarily and almost instantly change the shape of the crystalline lens so that the image of an object only a few inches distant, or an object miles away, may be distinctly focused upon the retina.

574. Visual Judgments. The eye has been likened to a camera, and so far as the physical principles involved are concerned, the likeness is very striking. The impressions received from the eye, however, are quite different from those given by a photograph, and for two reasons. In the first place a photograph gives a picture of an object in a given position, while the brain receives from the eye a composite impression. The difference between these two effects is very well shown in the comparison of a snapshot photograph of a horse in the act of running, with that of a simultaneous impression received from the eye, Fig. 564. And in the second place, our judgments of the position, size, and form of objects depend not only upon the images formed by the eye, but

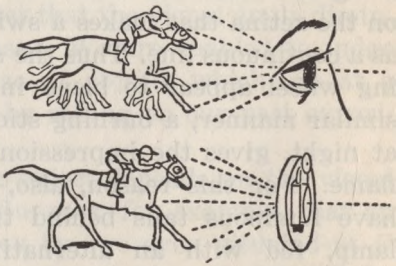


FIG. 564.—Images as formed by the eye and the camera lens

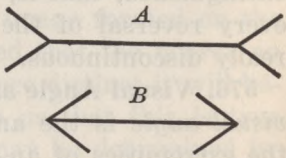


FIG. 565.—Horizontal lines *A* and *B* are of the same length

also upon our past experience. Sometimes our judgment is at fault, however, giving rise to optical illusions of various sorts. Thus in Fig. 565 the two horizontal lines are of exactly the same length. Line *A* appears to be longer than *B* because of the effect of the oblique lines at the ends.

575. Duration of Visual Impressions. The duration of a visual impression depends upon the sensitiveness of the retina and the intensity of the light. The duration of distinct vision is usually about one-sixteenth of a second. Distinct impressions, therefore, cannot be made upon the retina unless they succeed each other at intervals greater than that of the duration of the visual impressions for the given individual. It is this persistence of impressions on the retina that makes a swiftly moving object appear as a continuous line. Thus the spokes of a rapidly revolving wheel appear to blend into one another, and in a similar manner, a burning stick, whirled rapidly around at night, gives the impression of a continuous circle of flame. For this reason, also, shooting stars appear to have luminous tails behind them, and an electric arc lamp, fed with an alternating current of frequency greater than 25 cycles per second, appears to give a continuous light. If the arc be photographed, however, on a rapidly moving plate, it will be found that the light is extinguished, that is, diminishes to zero intensity, with every reversal of the current, thus showing that it is really discontinuous.

576. Visual Angle and Apparent Size of Objects. The *visual angle* is the angle formed by rays passing from the extremities of an object and intersecting at the eye, Fig. 566. The size of this angle depends (a) upon the size of the object, and (b) upon its distance from the eye; the greater the distance the less the visual angle. Now the apparent size of an object depends upon the

size of the visual angle, because this angle determines the size of the image formed upon the retina. Therefore the smaller the visual angle, the smaller will be the image on

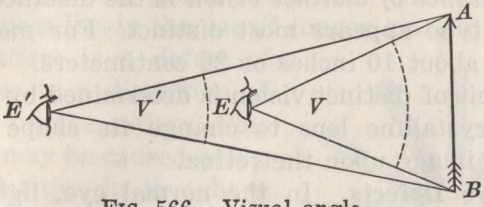


FIG. 566.—Visual angle

the retina, and consequently the smaller the apparent size of the object. If one looks along a railroad track, the rails appear to approach each other and the ties to grow shorter, due to the fact that the visual angle diminises as the distance increases. We are enabled to judge of the real height and size of objects which are at a considerable distance only by calling on our past experiences and judgments.

The apparent size of an object depends on the visual angle which it subtends. Suppose, for example, that an object *AB*, Fig. 566, is seen by two eyes, situated at *E* and *E′* respectively. The apparent size of *AB* will be greater for eye *E* than for eye *E′* because the visual angle *V* is greater than visual angle *V′*.

577. Distance of Distinct Vision. Distinctness of vision depends upon the size of the image formed on the retina; therefore it may be assumed that the nearer an object is brought to the eye, the more distinct it will become. Common experience teaches us that this is true only within certain limits, which may be determined by experiment. If the page of a book be held off at arm's length, the printed matter will not be distinct enough for most people to read with ease. Upon bringing the book gradually nearer, however, the type will become

more distinct as the page approaches, up to a certain point, from which position it will gradually grow dim again.

The *distance of distinct vision* is the distance at which ordinary type appears most distinct. For most normal eyes it is about 10 inches or 25 centimeters.

The limit of distinct vision is determined by the power of the crystalline lens to change its shape and thus focus the image upon the retina.

578. **Eye Defects.** In the normal eye, light from a given object is focused exactly upon the retina. The

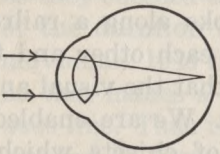


FIG. 567

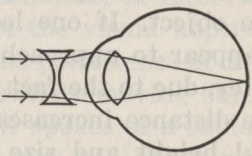


FIG. 568

Nearsighted eye, and correction by use of concave glasses

three most common defects of the eye are (a) nearsightedness, (b) farsightedness, and (c) astigmatism.

The *nearsighted* eye is one in which the rays from the crystalline lens come to a focus before reaching the retina, Fig. 567. Persons having such eyes are said to be nearsighted, because they have to bring the object

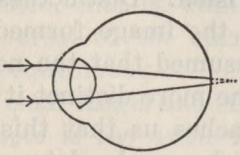


FIG. 569

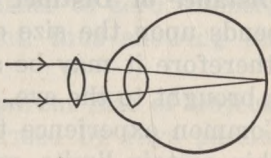


FIG. 570

Farsighted eye and correction by use of convex glasses

very near the eye in order to see it distinctly. Nearsightedness is remedied by the use of concave glasses,

which diverge the rays and thus throw the image farther back upon the retina, Fig. 568.

The *farsighted eye* is one in which the image tends to form beyond the retina, Fig. 569. The remedy for farsightedness lies in the use of convex glasses, Fig. 570.

Astigmatism is a defect of the eye due to a distortion of the image on the retina. This distortion may be caused either by irregularities in the shape of the eyeball or a lack of symmetry in the crystalline lens. The defect of astigmatism may be detected by looking at a series of radiating lines, Fig. 571. If the eye be astigmatic, the radiating lines will not all be seen with equal distinctness. Astigmatism may be corrected by the use of specially ground lenses which compensate for the defects of curvature in the eye.

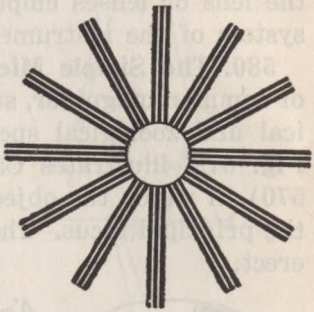


FIG. 571
Test for astigmatism

Lens formulas for glasses suitable for the correction of nearsightedness and farsightedness may be written as follows: (a) *Nearsighted eye.* Concave lens. Then $1/d - 1/d' = -1/f$ in which f is the focal length of the required eye-glass lens, d = distance of distinct vision (25 cm = 10 in.), and d' = distance that print has to be held from the eye to read without glasses. (b) *Farsighted eye.* Convex lens, object between the principal focus F and the lens. Then $1/d - 1/d' = 1/f$. The problem in each case is to find a lens of focal length such that an object (the printed page) at the distance of distinct vision (25 cm = 10 in.) shall appear to the eye as if at d', the distance of vision without glasses.

579. Magnifying Power. The *magnifying power* of an instrument, or *magnification* as it is sometimes called, is the ratio of the apparent size of an object as seen through the instrument to its real size; that is,

$$magnification = \frac{size\ of\ image}{size\ of\ object}$$

Since it is not always convenient or possible to measure the apparent size of an object as seen through the instrument, it is the usual practice to express magnifying power in terms of certain constants, such as the distance of distinct vision, the focal length or lengths of the lens or lenses employed, and the length of the lens system of the instrument.

580. The Simple Microscope. The *simple microscope* or common magnifier, such as is used in examining botanical and zoological specimens and other small objects, Fig. 572, illustrates Case 2 of the common lens (Art. 570) in which the object is placed between the lens and the principal focus. The image is virtual, magnified, and erect.

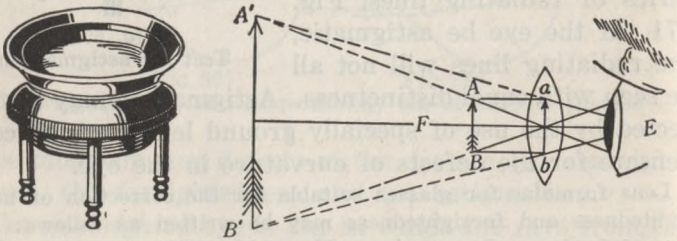


FIG. 572.—Simple microscope

The supporting legs of the instrument must be short enough so that an object placed on the table or desk upon which they rest lies between the principal focus and the lens. In other words, the length of the legs must be less than the focal length of the lens.

The magnifying power of the simple microscope is

$$\textit{magnification} = \frac{\textit{distance of distinct vision}}{\textit{focal length}} = \frac{10\,in.}{f} = \frac{25\,cm}{f}$$

in which $f =$ focal length of the lens in inches or centimeters.

581. The Compound Microscope. The *compound microscope* is a relatively high-power magnifying instru-

ment, Fig. 573, having two systems of lenses. The lower lens system which is near the object to be examined is called the objective, and the upper system is the eyepiece.

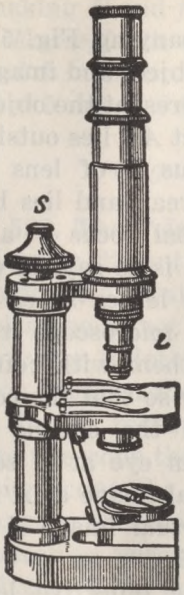


FIG. 573
Compound microscope

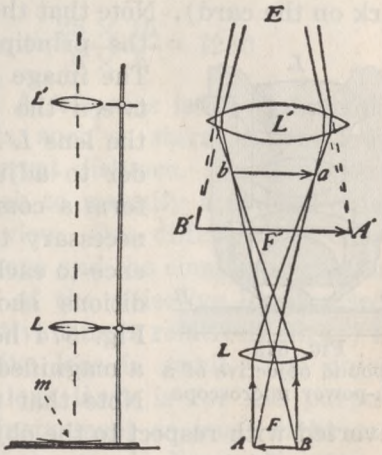


FIG. 574
Demonstration compound microscope

A compound microscope suitable for laboratory purposes may readily be constructed by mounting two convex lenses on a standard, as shown in Fig. 574. The objective lens L should have a focal length of about 5 centimeters; the eyepiece lens L', a focal length of about 25 centimeters. Place the lenses in position on the support, the lens having the short focal length being at L. Place a card bearing printed matter or a small pencil mark upon it below the lens, at m. Lower lens L until the object (mark on the card) lies just a little outside of the principal focus. Now move lens L' up and down until a position is found such that a clear and magnified

image of the object is seen when looking vertically downward along the dotted line.

A study of the outline sketch accompanying Fig. 574 will make clear the relative position of object and images in the compound microscope. Let AB represent the object (mark on the card). Note that the object AB lies outside the principal focus F of lens L. The image ab is real, and lies between the principal focus F and the lens L'. That is to say, in order to adjust the lenses so as to form a compound microscope it is necessary to set them with reference to each other so that the conditions shown in the outline in Fig. 574 hold. An eye at E sees a magnified virtual image at $A'B'$. Note that the virtual image $A'B'$ is inverted with respect to the object AB.

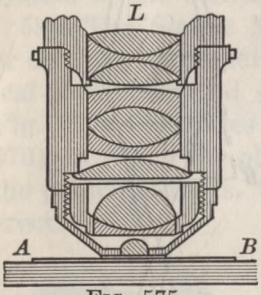


FIG. 575
Compound objective of a high-power microscope

In some high-power microscopes the objective lens system may consist of a combination of as many as ten lenses, as shown in Fig. 575.

The magnifying power of the compound microscope is

$$\textit{magnification} = \frac{25 \times L}{F \times f}$$

in which L = length of the microscope tube, F = focal length of the objective lens O, and f = focal length of the eyepiece lens E. When the constant 25 is used in the numerator, L, F, and f are expressed in centimeters. If L, F, and f are given in inches, then the number 10 should be substituted for the constant 25.

The equation for magnifying power tells us that there are two ways of increasing the magnification of a microscope, namely, by increasing the length L of the tube

and by using lenses of small focal length, that is, by making F and f as small as possible.

Example. In a given microscope $F = 5$ mm, $f = 6$ mm, and $L = 15$ cm. Find the magnifying power. *Solution*: $F = 5$ mm $= 0.5$ cm, and $f = 6$ mm $= 0.6$ cm. Then

$$\textit{magnification} = \frac{25 \times 15}{.5 \times .6} = 1250$$

582. The Telescope. A *telescope* is an instrument for viewing distant objects, such as stars or bodies on the earth at a relatively great distance. The function of a telescope is not so much to magnify a distant object as to bring it into clear view. One difference in construction between the telescope and the compound microscope is in the relative size of the objective lenses used. In the telescope the objective lens is relatively large; in the microscope the objective lens is small. The object of making the objective lens large is for the purpose of collecting within the instrument as much light as possible so as to admit of the required magnification without too much loss in brightness. In the Yerkes telescope, Yerkes Observatory, University of Chicago, the objective lens is 40 inches in diameter. This great lens is mounted in a tube 62 feet in length.

The principle of the telescope may be demonstrated in the laboratory by mounting two lenses, as shown in Fig.

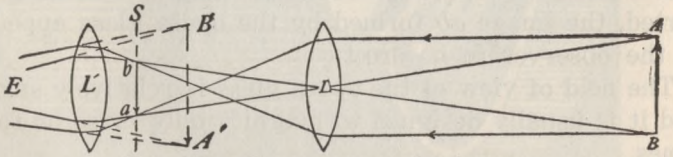


FIG. 576.—Lens system of telescope

576. The focal length of L, the objective lens, should be about 50 centimeters, and that of L' about 10 centi-

meters. Let the object AB be a candle or better an electric light. Mount lens L and find the position of the real image ab by means of a paper screen S. Now mount lens L' in such a position that the image ab lies between lens L' and its principal focus. An eye at E will see a virtual, magnified image of AB at $A'B'$.

The magnifying power of the telescope is

$$magnification = \frac{F}{f}$$

in which F = focal length of the objective lens O and f = focal length of the eyepiece lens E.

583. The Opera Glass and Prism Binocular. The *opera glass*, Fig. 577, is a short telescope, the eyepiece E of which is a concave lens. The focal length of E is the

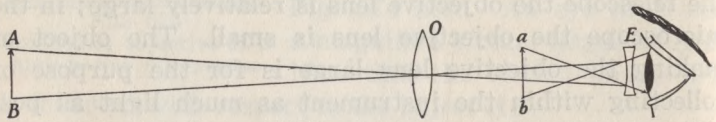
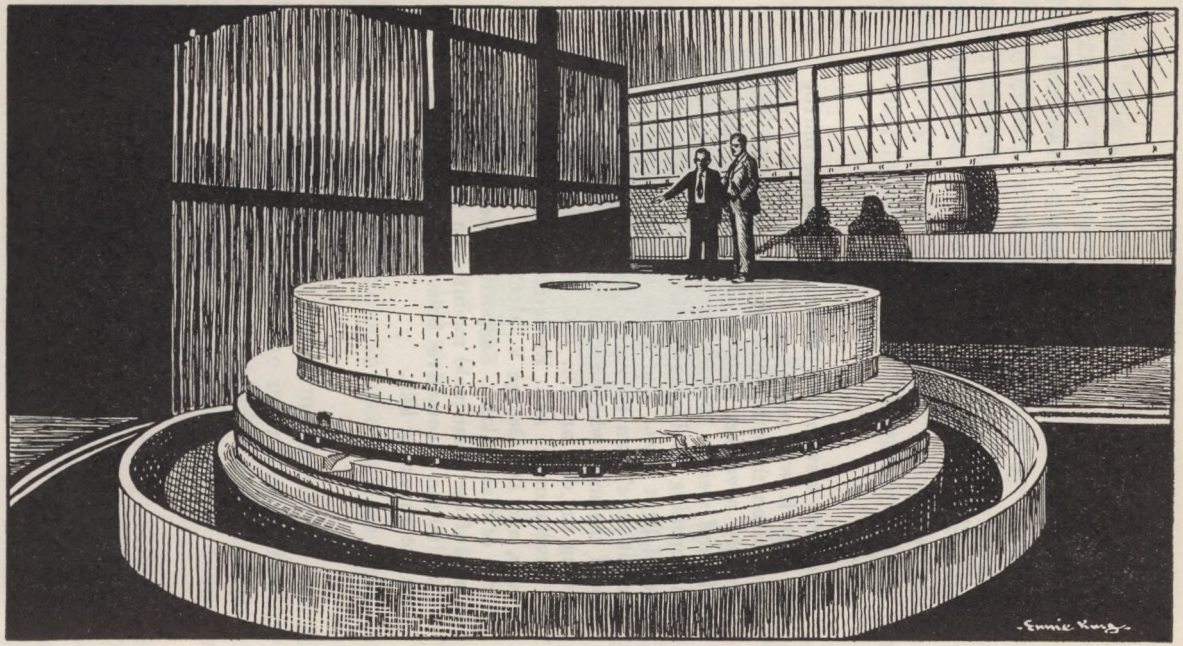


FIG. 577.—Sectional view of opera glass

same as the focal length of the eye. The effect of this is that the eyepiece of the instrument neutralizes the lens of the eye, and as a result the objective O forms an inverted image directly upon the retina. Since the image formed on the retina by the unaided eyes is always inverted, the image ab formed by the opera glass appears to the observer to be erect.

The field of view of the opera glass is relatively small, and it is usually designed to magnify only three or four times.

In recent years there has come into use the *prism binocular*, Fig. 578, which combines the compactness of the opera glass with the wide field of view of the tele-



THE LARGEST SOLID PIECE OF GLASS EVER CAST BY MAN

The first view of the 200-inch telescope mirror disc after it was poured at Corning, New York, is shown on the other side of this page. The disc is nearly seventeen feet in diameter, 27 inches thick, and weighs 20 tons. In the center is a circular opening 40 inches in diameter for the passage of light rays. The giant size of the world's greatest "eye" can be judged by the comparative size of the men who directed its production. Examining the disc are Dr. J. C. Hostetter, Director of Development and Research, Corning Glass Works, and Dr. George V. McCauley, physicist in charge of disc making. Supporting the disc is the bottom of the great mold mounted on the carriage of the 60-ton locomotive hoist which moved it from the annealing kiln.

scope. The arrangement of the objective and eyepiece is practically the same as in the astronomical telescope. The necessary distance between the objective lens and the eyepiece, however, is obtained by having the light travel the length of the tube three times in passing from the objective to the eyepiece. This is accomplished by reflection by means of a system of prisms within the tube of the instrument. The light from the object enters at

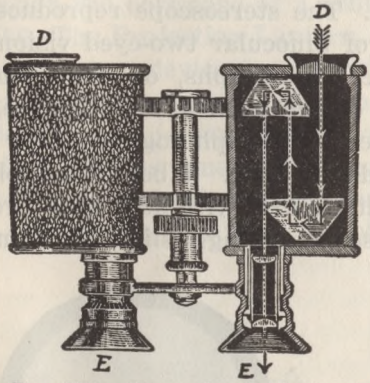


FIG. 578.—Binocular field glass

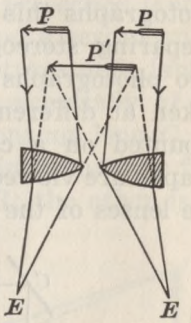


FIG. 579
Stereoscope

D and after four reflections in each tube emerges through the eyepiece at *E*.

The reflecting prisms perform the double function of diminishing the length of the instrument, thus giving it compactness, and also of reinverting the image so that the real image which is formed at the focus of the eyepiece is erect.

584. The Stereoscope. Nearly every child in school nowadays is familiar with the *stereoscope*, Fig. 579, an instrument for viewing photographs, which are mounted in pairs, and which are held in place in a receiving frame. The function of the stereoscope is first to give a

single slightly magnified image of the object, and second, and primarily, to give an image the characteristic appearance of *depth* as well as of area.

When an object is seen with both eyes, the image formed on the two retinas differ slightly, because of the fact that the two eyes are viewing the object from slightly different angles. It is this difference in the two images which gives a landscape viewed with two eyes the appearance of depth, or solidity, which is absent when one eye is closed. The stereoscope reproduces in photographs this effect of binocular two-eyed vision. In preparing stereographic photographs, or *stereographs*, two photographs of an object, such as a landscape, are taken at different angles. These photographs are then mounted on a card side by side. When these photographs are viewed simultaneously with both eyes through the lenses of the stereoscope, a single slightly magnified

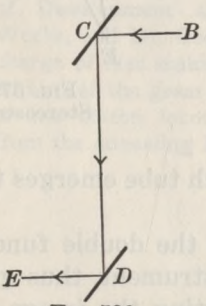


FIG. 580

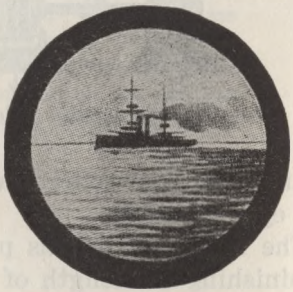


FIG. 581

The periscope and field of view

image is seen which presents the same characteristic appearance of depth as when viewed under normal conditions with both eyes.

585. The Periscope. A *periscope* is a jointed telescope which possesses two or more reflecting surfaces, and which enables one in a trench or in a submarine to see

around a corner or over an intervening object. The principle of the periscope may be illustrated in a very simple way by mounting two mirrors, as shown in Fig. 580. Light from B is reflected at C and D to the eyes at E. The type of periscope which is used on submarines contains a system of lenses which gives a magnification necessary to make ships appear at their true distance from their observer.

In Fig. 581 there is shown the field of vision as seen through the periscope of a submarine.

586. The Projection Lantern. The *projection lantern*, Fig. 582, is a device for throwing upon a screen an enlarged image of an object such as pictures on a lantern slide. The projection device consists essentially of two optical systems, one called the condensing lens C, the function of which is to condense the divergent rays from the electric light L upon the slide AB; the other is the

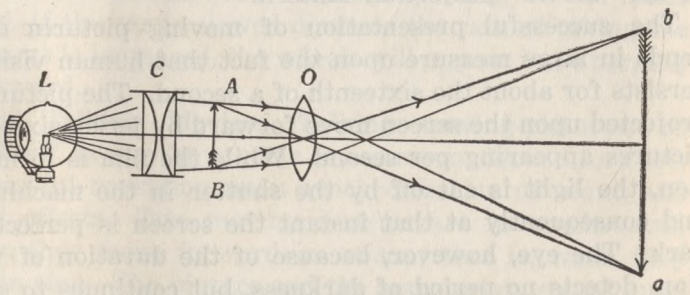


FIG. 582.—Projection lantern

objective lens O, the function of which is to focus the light upon the screen. The image is real and inverted.

The distance of the slide AB from the objective lens O is called the object distance, and the distance of the image on the screen from objective lens O is called the image distance. In order to focus the lantern so as to bring out clearly the picture upon the screen, the lens O

is moved back and forth until a sharp image is produced upon the screen.

Formerly the electric arc was used as the source of light in projection lanterns. In recent years, however, these have been almost entirely displaced by the high power gas filled incandescent lamp. A spherical reflecting mirror is placed immediately behind the bulb to increase the intensity of the light upon the condensing lens.

587. Moving Pictures. In the preparation of moving pictures, a camera is employed which takes a series of snapshots upon a film which moves forward by jerks. While the film is being jerked forward, a shutter, operated mechanically, momentarily covers the lens. The pictures are taken between jerks, at which time the shutter is opened, exposing the lens. The usual length of film per reel is 1,000 feet; this contains 16,000 pictures. From the negative reel of film positive films are printed for use in the "movie" projection lantern.

The successful presentation of moving pictures depends in large measure upon the fact that human vision persists for about the sixteenth of a second. The pictures projected upon the screen move forward by jerks, sixteen pictures appearing per second. While the film is in motion, the light is cut off by the shutter in the machine, and consequently at that instant the screen is perfectly dark. The eye, however, because of the duration of vision, detects no period of darkness, but continues to see the picture which was visible the instant before. Before the vision of this picture dies out, another flashes into view, and so on. It thus appears that the so-called moving pictures do not move at all so far as retinal vision is concerned. To be sure the pictures on the reel move forward at the rate of sixteen per second, but the series of pictures which the eye perceives are for the instant stationary, the gap between any one and the next suc-

ceeding picture being filled by the power of the retina to retain visual impressions.

The fact that intermittent exposures are made in the taking of the pictures sometimes gives rise to curious optical illusions. For example, the wheels of an automobile which is in motion sometimes seem to cease rotating and sometimes they even seem to rotate backward. Let us suppose that an automobile wheel has twelve spokes and that it is making three revolutions per second. This means that 36 spokes will appear to pass a given point per second. If now the pictures are taken at the rate of 36 per second, the wheels will appear to be stationary in precisely the same way that two trains running at the same speed in the same direction on parallel tracks appear to be stationary with respect to each other. If now the speed of the automobile is slightly increased, or the rate of taking the pictures decreased, the wheels will appear to turn forward very slowly; and, on the other hand, if the car is slowed down or the rate of taking the pictures is increased, the wheels will appear to be turning backward.

588. **Talking Movies.** Talking moving-pictures, or "talkies" as they have come to be known, depend for their successful operation upon synchronization of sound and light effects. Two moving systems are in synchronism when they move forward in unison; that is, when they are in step. The picture and the appropriate sound must appear simultaneously. To accomplish this two general methods are employed. These are (a) the disc method, known as the Vitaphone, and (b) the film method represented by the Movietone and the Photophone. The Movietone and Photophone make use of practically the same principle, there being only a slight mechanical difference in the set up.

In the disc or Vitaphone method the sound record is on

a disc, which is similar to a phonograph record, except that it is much larger and is designed to run at about half the standard speed. Synchronism between sound and picture is secured by having the projector and the turntable carrying the disc both driven by the same motor.

In the film method, both picture and sound records are photographed side by side on the film, Fig. 583. Sound records consist of a narrow band called the "sound track" which runs down the side of the film. The sound track is crossed by microscopic lines, the spacing of which at any point depends upon the pitch of the sound which was recorded. The difference in density of the lines depends upon the loudness of the sound; that is, the greater the contrast between the light and dark lines on the sound track, the louder the sound. In ordinary moving picture projection, the film is usually shown at a faster speed than it was taken. This cannot be done with a sound film because the pitch would be changed and this would cause the voice or music to be distorted and spoiled. All synchronized subjects therefore have to be shown at exactly the same speed at which they were made, about 90 feet per minute.



FIG. 583.—Section of a "talkie" film

When the movietone machine is in operation the sound film, which passes over a sprocket wheel, moves along through the reproducing chamber at a constant speed. A narrow, bright beam of light is focused on the sound

track of the film. The light which has passed through the moving film varies in intensity according to the variation of the lines recorded on the sound track. This light falls upon a photo-electric cell (Art. 480) which produces a small electric current whose vibrations correspond to the light and therefore to the sound which was recorded. That is, the mild current from the photo-electric cell passes along to one or more tube amplifiers, similar in appearance to those used in audio frequency stages of radio sets. The current from the amplifiers is converted into sound by means of sound projectors which are placed immediately behind the screen so that a perfect illusion that the sound is coming from the speaker or other source seen on the screen is obtained in all parts of the house.

Color and Dispersion

589. Introductory. Light, we have learned, is a vibration which is capable of affecting the organs of sight. *Color,* on the other hand, is a sensation; it depends upon the wave length and frequency of the light which falls upon the retina. The color of red is due to light of relatively long wave length; violet is due to light of short wave length. We cannot properly speak of red light or violet light, since the waves which give rise to these sensations (colors) are in themselves colorless. When light waves making 395 trillion vibrations per second fall upon the retina, they give rise to a sensation which we have been taught to call red; and likewise, light waves making 760 trillion vibrations per second give rise to a sensation which we call violet. In a similar manner we can account for all the intermediate color sensations.

It must be noted here that light is the cause of color and that the two important conditions for the production of color are (a) the presence of light of a given wave

length, and (b) a sensorium (nerves and nerve centers) to receive the impression. If there were, therefore, no eye to receive the impressions of the light waves, there would be no color.

Color, like the pitch of a musical tone in sound, depends upon the frequency, that is, upon the number of vibrations per second. The human ear is capable of perceiving musical notes covering a frequency range of about ten octaves, that is, from 30 vibrations per second to about 30,000. The human eye, on the other hand, is capable of perceiving colors over a frequency range of only one octave, that is, from 395 trillion vibrations per second to about 760 trillion.

590. The Solar Spectrum. If a beam of white light from the sun or an electric arc be allowed to pass

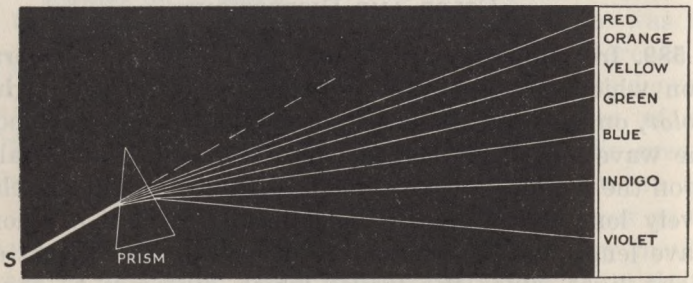


FIG. 584.—Solar spectrum

through a prism and fall upon a white screen, Fig. 584, it may be observed that the band of white light is broken up into the seven colors of the solar spectrum, as follows: *red, orange, yellow, green, blue, indigo,* and *violet.*

Red is refracted (bent out of its course) least and violet most. The rule that applies to refraction with reference to wave length is this: *The greater the wave length the less the refraction.* Since each color band of the solar spectrum is relatively wide, it follows that the

wave length for any color will depend upon what part of the color band is selected. Taking the *A* line of the red band and the *H* line of violet (See Fig. 586, and Table XX, Appendix) the respective wave length and frequency of these colors are

	Wave Length in mm.	*No. Vibrations per second*
Red	.000760	395,000,000,000,000
Violet	.000397	760,000,000,000,000

A *spectrum* is the series of colors of which the light from a given source is composed. This breaking up of white into its prismatic colors is called *dispersion.*

591. The Rainbow. One of the most familiar and striking examples of dispersion is that seen in the rainbow. In order to see a rainbow three conditions are necessary: (a) There must be drops of water in a cloud or mist; (b) the sun must be shining on the cloud; and (c) the eye of the observer must be in such a position as to receive the refracted light from the raindrop; that is, the observer must stand with his back to the sun.

There are two rainbows, the primary and the secondary. The *primary bow,* which is the one usually observed, is much the brighter of the two and lies inside (below) the secondary. In the primary bow the red is on the outside and the violet on the inside.

The *secondary bow* is the fainter of the two, and in fact is not usually seen except under very favorable conditions. Its colors are reversed as compared with those of the primary, the violet being on the outside and the red on the inside.

In Fig. 585 we have given a more or less ideal sketch illustrating the relation of the primary to the secondary bow with reference to the number of reflections and refractions in each. The margin of each bow is selected

for illustration. In order to see both bows, the eye E must be in such a position with respect to the sun S that the light refracted and dispersed by the drops will make

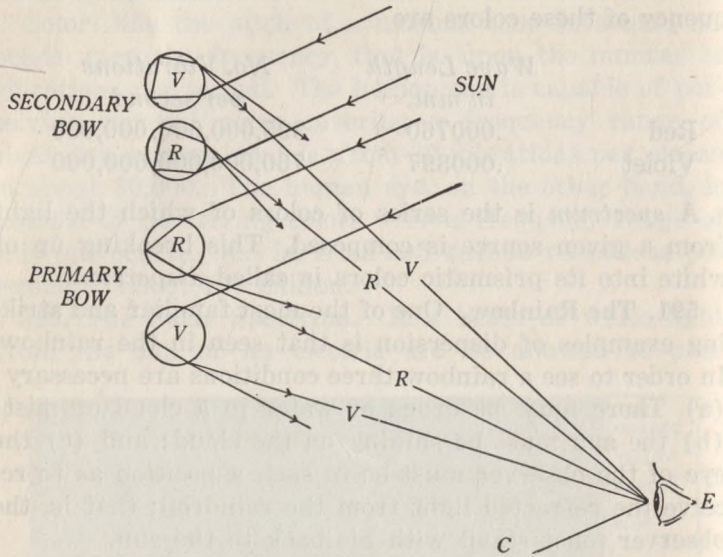


FIG. 585.—Illustrating the formation of primary and secondary rainbows

a definite angle with the line of direction (EC) of the sun's rays. This angle depends upon the index of refraction of the drop and the color seen by the eye. This angle varies from color to color since the index of refraction of the different colors of the spectrum vary. The angle REC for red of the primary bow is 42°, that of violet, VEC, 40°; for the secondary bow the angles are, for violet 54°, for red 51°. The reason that the rainbow is circular, then, is that the color from only those drops which makes a definite angle with EC are seen by the eye. In other words, the eye is at the ver-

tex of a cone, a portion of the base of which is a circle forming the rainbow.

It will be observed that light forming a primary bow suffers *two* refractions and *one* internal reflection; the secondary bow, *two* refractions and *two* reflections. Now since some light is lost at each reflecting surface, it is easy to understand why the primary bow is the brighter of the two.

592. Kinds of Spectra. There are three kinds of spectra, as shown in Figs. 586, 587, 588, namely, discontinuous or dark line, continuous, and bright line.

A *discontinuous spectrum,* Fig. 586, is one in which dark lines appear at intervals across the color band of which the spectrum is composed. Such a spectrum is formed when the light from an *incandescent solid passes through that of an incandescent gas.* The solar spectrum is a dark-line spectrum. It is true that it appears to the naked eye to be continuous. If, however, the solar spectrum be examined by means of a spectroscope, it will be found that many dark lines appear in it, as shown in Fig. 586. These are called *Fraunhofer lines,* after Joseph Fraunhofer, a German scientist, who was one of the first to count and describe them. The fact that the sun's spectrum is a dark-line spectrum tells us that its light comes from an incandescent solid, liquid, or gas under high pressure surrounded by an incandescent gas under relatively low pressure.

A *continuous spectrum,* Fig. 587, is one in which the colors blend gradually from one into the other without any break. The spectrum from an electric lamp is continuous, the colors grading continuously from red to violet. An *incandescent solid* gives a continuous spectrum. The spectrum from a gas flame, or a candle flame, or the flame of a kerosene lamp is in all cases continuous, the luminous properties of such flames being due to the

red hot solid particles which they contain. Likewise a platinum wire or any other metal heated to incandescence gives a continuous spectrum.

A *bright-line spectrum,* Fig. 588, is one consisting of one or more bright lines. A bright-line spectrum is formed from an *incandescent gas.* Thus if a piece of sodium be placed in the flame of a Bunsen burner and the yellow light examined by means of a spectroscope, a bright yellow band will appear. This is a bright-line spectrum. So long as a substance remains in the state of an incandescent solid, it gives a continuous spectrum; the moment, however, that it becomes gasified and incandescent, it gives a bright-line spectrum.

593. Spectrum Analysis. If any substance in the condition of an incandescent gas be examined by means of a spectroscope, Fig. 589, it is possible to determine from the nature of its spectra many facts regarding the character of the substance. For example, suppose that we wish to determine if there is any sodium in a given inorganic substance. Dip a platinum wire into the substance and then place it in the colorless part of a Bunsen flame; the salt will be volatilized, and if any sodium be present, a distinct yellow color will be imparted to the flame. If this be examined by means of a *spectroscope,* the presence of even the most minute trace of sodium will reveal itself by the appearance of a yellow band, such as is shown in Fig. 588.

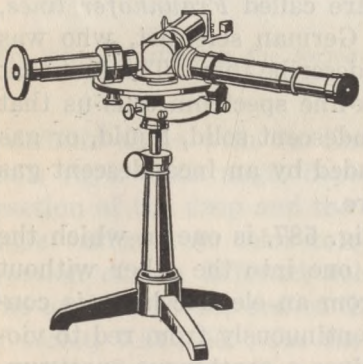


FIG. 589.—Spectroscope

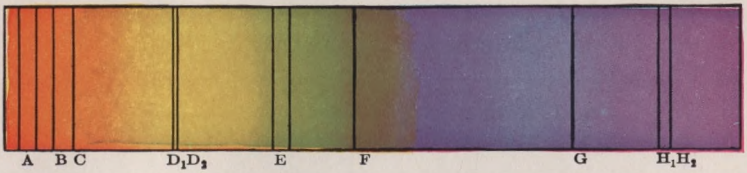


FIG. 586.—Discontinuous or Dark Line Spectrum. Crossing this spectrum there are many dark lines which are called Fraunhofer lines. These lines, however, are not visible under ordinary conditions without the aid of a spectroscope.



FIG. 587.—Continuous Spectrum. The colors of a continuous spectrum blend one into the other without any dark lines occurring, as in the case of the spectrum of Fig. 586. A continuous spectrum is due to incandescent solids, such as the carbons of the electric arc, or the glowing filament of an incandescent lamp.



FIG. 588.—Bright Line Spectrum. A bright line spectrum is formed by the light from an incandescent gas. Each metal, for example, when in a volatile condition and heated to incandescence gives rise to its own characteristic bright line spectrum, the above being that of sodium.

Thus it is possible to determine the presence of less than one ten-thousandth of a gram of sodium.

Also, when the spectrum of the sun and the stars are examined and we find present the lines that are characteristic of sodium, iron, magnesium, and other metals, we take for granted that these metals are present in the heavenly bodies from which the light comes.

594. The Invisible Spectrum. The visible spectrum is that which is included between red and violet, about one octave of color; that is, from red making 395 trillion to violet which makes 760 trillion vibrations per second. The visible spectrum, however, is not the limit of the dispersion of ether waves which may be secured by a prism. Beyond the red is a series of longer waves which manifest themselves to our senses in the form of heat. Below the violet there are very short waves which are capable of great chemical activity. The invisible rays beyond the red are called the *infra-red*; those beyond the violet are called the *ultra-violet.*

595. Analysis and Synthesis of White Light. The white light of the sun may be considered as made up of many colors, the most striking of which are the seven so-called spectral colors which appear when white light is dispersed by a prism. The question naturally arises:

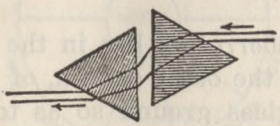


FIG. 590
Analysis and synthesis of white light

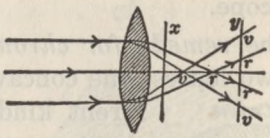


FIG. 591
Chromatic aberration

Is it possible to combine these seven colors so that they may give us again white light? Sir Isaac Newton asked himself this very question and was one of the first to

answer it by means of an experiment. He first passed a beam of white light through a prism, causing it to become dispersed; he then passed the dispersed beam through a second prism, inverted with respect to the first, Fig. 590, and found that the colors of the spectrum were combined by the second prism into a band of pure white light.

The breaking up of white light into its spectral colors is called analysis of white light; the combining of these colors is called synthesis.

596. Chromatic Aberration. When white light passes through a lens, two things occur: (a) it is refracted and (b) it is to a certain extent dispersed; that is, it is broken up into prismatic colors. This dispersion of white light which occurs in a lens is called *chromatic aberration* and is illustrated in Fig. 591. The violet rays, being the most refrangible, come to a focus nearer the lens than do those of the red. If, therefore, a white card be placed in the light at x there may be observed around the outer edge a fringe of red; if the card be placed at y, there will be seen around the outer edge a fringe of violet. This fringe of color constitutes a very serious defect in the image, especially where it is desired to produce a clear image, as in the case of the microscope.

The *remedy for chromatic aberration* lies in the use of two lenses, one concave and the other convex, of different kinds of glass ground so as to fit closely one upon the other. In Fig. 592 there is shown an *achromatic lens* consisting of a bi-convex lens of crown glass and a plano-concave lens of flint glass. The light passing through such a lens is brought to a focus without being dispersed. Lenses of

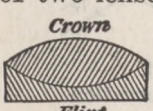


FIG. 592
Achromatic lens

this type are used in the objective of the compound microscope.

597. Mixing Colors. Since color is a sensation, it follows that the mixing of colors is nothing more than the mixing of sensations. If a color disc, sometimes known as Newton's disc, Fig. 593, be rapidly rotated, the eye will receive a series of impressions, the resultant sensation of which will be caused by the mixing of the several sensations due to the different colors on the disc. Thus if the disc be rotated, the eye will no longer see yellow, blue, green, etc., but will see in their stead gray or white, or whatever the resultant color may be.

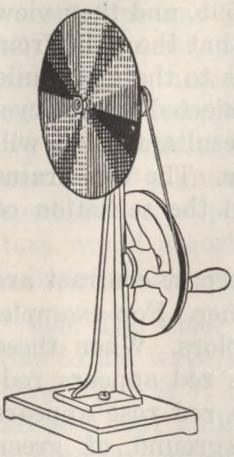


FIG. 593.—Color disc

598. Complementary Colors. Any two colors which when mixed produce the sensation of white are called

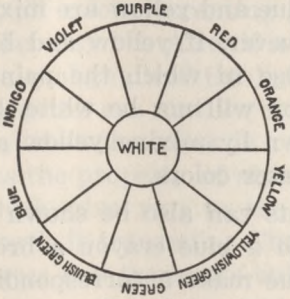


FIG. 594
Complementary color disc

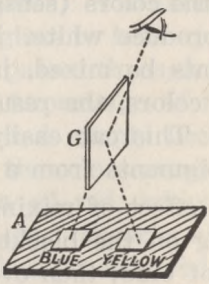


FIG. 595
A method of mixing colors

complementary colors. A complementary color disc is shown in Fig. 594. Those colors which are opposite each other on the disc will, on being mixed, produce white.

Thus yellow and dark blue are complementary, as may be shown by the following very simple experiment. Place two small pieces of paper, one yellow and the other blue, a few inches apart on the table, Fig. 595, and then view them by means of a piece of glass so that the light from the blue piece passes through the glass to the eye, while the light from the yellow paper is reflected to the eye. The sensation produced, that is, the resultant color, will be neither yellow nor blue, but white. The two trains of light waves unite and give rise to the sensation of white.

Those colors which produce the sharpest contrast are in general complementary to each other. For example, red and green are complementary colors. When these colors are brought near together, the red appears redder and the green greener. Thus the red rose appears redder when seen against the background of green leaves than it otherwise would. Blue also appears bluer when placed adjacent to a yellow field.

599. Mixing Pigments. We have just learned that when the colors (sensations) blue and yellow are mixed, they produce white. Now, however, if yellow and blue pigments be mixed, in the sense in which the painter mixes colors, the resulting color will not be white, but green. This can easily be shown by mixing yellow and blue pigments from a set of water colors.

The effect of mixing pigments can also be shown by making on the blackboard with a blue crayon a broad band of blue; then over the blue make a corresponding band of yellow. The mixture of blue and yellow crayon will give a color of greenish tinge.

Mixing pigments then produces an entirely different effect from the mixing of corresponding colors. Yellow and blue colors, to repeat, produce white; yellow and

blue pigments produce green. The reason that the mixing of blue and yellow pigments produces green is because the blue pigment absorbs all the colors of white light except blue and green, and the yellow pigment absorbs all but the red, yellow, and green. Now when the two are mixed, there results a pigment that absorbs every color but green, which is reflected. In a similar manner we may explain the resulting color due to mixing any number of pigments. If, for example, we should take a combination of pigments of such a nature that the mixture would absorb all the colors of white light, the resulting mixture would be black.

600. The Color of Bodies. The color of any body is due to the wave length of the light which it *reflects,* or *transmits.* If it reflects light having a wave length of about 0.0007 millimeter, it gives rise to the color of red, and so on throughout the color scale. We often speak of an object as being painted a given color, as, for example, red. What is really done is to put upon the object a pigment which has the property of absorbing the light of all colors excepting red, or green, or whatever the color may be. Also, if a piece of glass be held up before the sunlight, and the light which passes through it falls upon the eye and we receive the impression of red, we say that the glass is red. It is red simply because it has the property of absorbing all the colors except those which give rise to the sensation of red, which wave length it transmits. That is, the glass absorbs all other waves and transmits only those that produce red.

This all means that the color of any object is not something that resides within the object, but is a sensation having its seat within the brain of the observer. The red of the rose, for example, is not in the rose, but in the sensorium of the observer. The rose simply has

the property of reflecting those waves which give rise to the sensation of red.

601. Colors Due to Interference of Light Waves. If a soap bubble be observed in the sunlight, brilliant colors will be seen shifting rapidly across the surface of the bubble. These colors are caused by what is known as *interference.* Since the light is a wave motion, it follows that two waves may interfere in somewhat the same way that two sound waves interfere. It is possible, therefore, that two light waves may be superimposed one upon the other in opposite phase so that one may exactly neutralize the other, thus producing interference effects; also two light waves may be superimposed in the same phase producing re-enforcement, in which case the intensity of the resultant wave is greater than that of either of the waves which produced it. Interference of light may be demonstrated by means of the following experiment. Fasten two pieces of plate glass together by means of rubber bands, the lower ends of the two pieces being separated by a piece of tissue paper, so that there is a thin wedge of air between the plates. Wrap a piece of asbestos which has been soaked in a strong solution of common salt (NaCl) around the tube of a Bunsen burner so as to touch the flame, as shown in Fig. 596. The flame will be colored a bright yellow by the sodium in the salt. When the eye looks at the reflection of the flame from the glass surfaces, a series of fine black and yellow lines will be seen to cross the plate. These are interference bands.

In order to explain interference, let us consider Fig. 597 which shows an enlarged section of the glass plates with the wedge of air between them. Let the yellow sodium light come from the right in a series of transverse waves, which may be represented by the wavy lines.

This light is in part transmitted and in part reflected at each glass surface. We are interested, however, only in what happens at the interior faces *AB* and *CD* of the

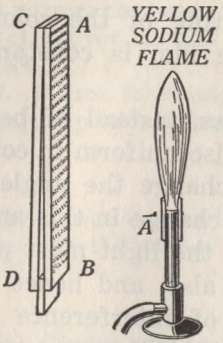


FIG. 596

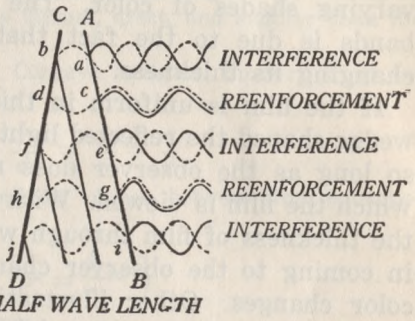


FIG. 597

Colors by interference

plates. The full lines represent the light waves reflected from the surface *AB* of the air wedge and the dotted lines represent the waves reflected from the surface *CD*. If the distance from *a* to *b* is such as to make one reflected wave just half a vibration behind the other in phase, they will interfere, producing a dark band. But if at another point the distance between the plates is such that the wave reflected at *d* coincides with the wave reflected at *c*, re-enforcement occurs. At this point there is a bright yellow line. If we select any two *consecutive* dark lines, we know that the *double* path between the plates at one line must be just one wave length longer than that at the other line; that is, the distance from *a* to *b* is one half wave length and the distance *ef* is one whole wave length. This gives us a method of measuring the wave length of light for any given color.

This explanation applies also to the brilliant bands of

color which occur in soap films. The light falling upon the soap film is not of one wave length, as in the case of the yellow sodium flame, but is composed of all the wave lengths of white light. Here we see bands of varying shades of color. The shifting of the colored bands is due to the fact that the film is constantly changing its thickness.

If the film is uniform in thickness, instead of being wedge shaped the reflected light is also uniform in color, so long as the observer does not change the angle at which the film is viewed. With any change in this angle the thickness of film through which the light must pass in coming to the observer changes also, and hence the color changes. Other illustrations of interference are seen in the brightly colored bands which appear when a film of oil spreads out over the surface of water; also the bright bands which appear due to cracks in ice, or in other crystalline substances, illustrate interference effects.

SUMMARY

1. There are two general classes of mirrors, namely, plane and spherical. Mirrors, however, are sometimes made cylindrical in shape, and sometimes paraboloid.

2. Images may be real or virtual. A real image is one that is formed by the actual focusing of the rays of light; a virtual image is one that is formed by the apparent focusing of the rays of light.

3. Images formed by plane mirrors are always virtual.

4. Images formed by spherical mirrors may be real or virtual, depending on the position of the object with respect to the mirror.

5. The defining terms associated with spherical mirrors are: vertex, center of curvature, principal axis, principal focus, focal length.

6. Images formed by spherical mirrors: (a) Concave mirror, object at a distance from the mirror greater than the principal focus: the image is real and inverted, its size depending on the distance of the object from the principal focus; when the ob-

ject is beyond the center of curvature, the image is smaller than the object, and when the object is between the center of curvature and the principal focus, the image is larger than the object.

(b) Concave mirror, object between the principal focus and the mirror: the image is virtual, erect, and larger than the object.

(c) Convex mirror: image virtual, erect, and smaller than the object.

7. Mirror formulas: (a) Concave spherical mirror, object beyond the principal focus,

$$\frac{1}{d} + \frac{1}{d'} = \frac{2}{r} = \frac{1}{f}$$

(b) Concave spherical mirror, object between the principal focus and the mirror,

$$\frac{1}{d} - \frac{1}{d'} = \frac{2}{r} = \frac{1}{f}$$

(c) Convex spherical mirror,

$$\frac{1}{d} - \frac{1}{d'} = -\frac{2}{r} = -\frac{1}{f}$$

8. Lenses are of two general classes, namely, convex and concave. A convex lens is one that is thicker at the middle than at the edge; a concave lens is one that is thinner at the middle than at the edge.

9. Defining terms for lenses: center of curvature, principal axis, optical center, principal focus, focal length.

10. Convex lenses are convergent, bringing rays to a focus; concave lenses are divergent, scattering the rays.

11. Images formed by lenses: (a) Convex lens, object beyond principal focus: the image is real and inverted, its size depending on the distance of the object from the principal focus.

(b) Convex lens, object between the principal focus and the lens: the image is virtual, erect, larger than the object.

(c) Concave lens: the image is virtual, erect, and smaller than the object.

12. Lens formulas. (a) Convex lens, object beyond the principal focus:

$$\frac{1}{d} + \frac{1}{d'} = \frac{1}{f}$$

(b) Convex lens, object between the principal focus and the lens:

$$\frac{1}{d} - \frac{1}{d'} = \frac{1}{f}$$

(c) Concave lens:

$$\frac{1}{d} - \frac{1}{d'} = -\frac{1}{f}$$

13. A nearsighted eye is one in which the rays passing through the crystalline lens come to a focus before reaching the retina. Remedy for nearsightedness: concave glasses.

14. A farsighted eye is one in which the rays passing through the crystalline lens tend to come to a focus behind the retina. Remedy for farsightedness: convex glasses.

15. The duration of a visual impression depends upon the sensitiveness of the retina and the intensity of the light. The duration of distinct vision is in general about one-sixteenth of a second.

16. The visual angle is the angle formed by rays passing from the extremities of an object and intersecting at the eye. The apparent size of an object depends upon the size of the visual angle.

17. The distance of distinct vision is the distance at which ordinary type appears most distinct. For the normal eye the distance of distinct vision = 10 in. = 25 cm.

18. Magnifying power of an optical instrument containing one or more lenses is the ratio of the size of the image to the size of the object; that is,

$$magnification = \frac{size\ of\ image}{size\ of\ object}$$

Magnification is usually expressed in terms of certain constants of the instrument.

19. Color is a sensation which depends upon the frequency of the light waves falling upon the retina. A frequency of 395 trillion light waves per second produces the sensation of red; a frequency of 760 trillion produces the sensation of violet.

20. The colors of the solar spectrum are in order from the lowest wave length to the highest: red, orange, yellow, green, blue, indigo, and violet. Red is refracted least; violet, the most.

21. There are three kinds of spectra: discontinuous or dark line, continuous, and bright line.

22. A discontinuous or dark line spectrum is one which is crossed by dark bands. A dark line spectrum is formed by light from an incandescent solid passing through an incandescent gas.

23. A continuous spectrum is one in which the colors blend one into the other without any break. A continuous spectrum is formed by light from an incandescent solid.

24. A bright line spectrum is one which consists of one or more bright lines. A bright line spectrum is formed by light from an incandescent gas.

25. Complementary colors are any two colors which when mixed produce the sensation of white.

26. The color of any body depends upon the wave length (and therefore the frequency) of the light which it reflects or transmits.

REVIEW EXERCISES

A

1. What is the difference between real and virtual images? What kind of images are always formed in plane mirrors?

2. Explain the formation of multiple images. Why, in certain reflecting instruments, are mirrors of polished metals used instead of glass mirrors?

3. In the case of multiple images formed in a plate glass mirror, which image in the series is the brightest? Why?

4. By means of a diagram show how two plane mirrors may be arranged so that a customer in a tailor shop may see his back.

5. If a person walks toward a plane mirror with a velocity of 8 ft. per second, what is the velocity with which the person and the image approach each other?

6. What is meant by the statement that images formed in plane mirrors are reversed?

7. Distinguish between concave and convex mirrors.

8. By means of a diagram of a spherical mirror, locate and explain the meaning of the following: vertex, center of curvature, principal axis, principal focus, focal length.

9. The distance from the vertex of a spherical mirror to its center of curvature is called the "radius of curvature." What is the relation between the focal length of a spherical mirror and its radius of curvature?

10. What is a convex lens? What is its effect on parallel rays of light? Name and illustrate by drawing the three classes of convex lenses.

11. What is a concave lens? What is its effect on parallel

rays of light? Name and illustrate the three classes of concave lenses.

12. By means of an outline drawing of a double-convex lens, locate and explain the meaning of the following: centers of curvature, principal axis, optical center, principal focus, focal length.

13. What are conjugate foci? Give an illustration.

14. Is it true that real images formed by both mirrors and lenses are always inverted?

15. A small object at a distance greater than the radius of curvature from a convex lens is moved toward the lens. What change occurs in the position and size of the image as the object approaches the principal focus?

16. Describe the camera. (a) Is the image formed on the film real or virtual? (b) Is it erect or inverted? (c) What is the use of the diaphragm? (d) the shutter?

17. Make a sketch to illustrate a cross section of the human eye. Name the different parts and explain the function of each.

18. Make a sketch to illustrate a nearsighted eye, and explain how this defect may be remedied.

19. Make a sketch to illustrate a farsighted eye, and explain how this defect may be remedied.

20. What is astigmatism, and how may its presence be determined?

21. Make a sketch to illustrate visual angle. What is the relation of the apparent size of an object to the visual angle?

22. What is the distance of distinct vision for the ordinary eye? What is the distance in your case? Is it the same for both eyes?

23. What is meant by "persistence of visual impression"? What is the average duration of distinct vision? What modern application has there been of this power of the eye to retain visual impression?

24. (a) What is magnifying power? A word printed on the page of a book is a quarter of an inch in length; as viewed through a lens, it appears to be half an inch in length. (b) What is the magnification in this case?

25. The image formed on the retina by the crystalline lens is inverted. How do you account for the fact that objects seen by the eye appear right side up?

26. Explain why an object seen through an opera glass is erect.

27. Account for the appearance of depth, which a double photograph (stereograph) possesses when viewed through a stereoscope.

28. Why are lantern slides inserted in the projection lantern upside down?

29. As a projection lantern is moved back from the screen, what change takes place in the size of the picture?

30. Why are moving pictures thrown on the screen at the rate of 16 per second?

31. Why is the so-called moving picture not really a moving picture?

32. (a) Define color, and name the colors of the solar spectrum. (b) Compare the vibration frequency of light producing the color red with that producing violet.

33. If the waves producing the sensation of red were all absorbed from sunlight, what color would objects that were formerly red appear to have?

34. Why does a blue dress look black in yellow light?

35. From which end of the solar spectrum do we get (a) the greatest thermal (heat) effects? (b) the greatest chemical effects?

36. Why does the photographer in developing pictures in the "dark room" use red light?

37. In what part of the sky (east or west) will a rainbow appear if it is formed (a) early in the morning? (b) late in the afternoon? (c) Why is a rainbow never seen in the middle of the day?

38. Suppose that a person were in a balloon above the clouds, what shape of rainbow might it be possible to see?

39. Explain the mixing of colors by means of Newton's disc.

40. What are complementary colors? Give some examples. How do complementary colors affect each other when placed side by side?

41. Why does a red rose appear redder when seen against a background of green leaves than when viewed against a background of another color?

42. Distinguish between mixing colors and mixing pigments.

43. (a) What color is produced when blue and yellow colors

are mixed? (b) What color is produced when blue and yellow pigments are mixed? Explain.

B

44. Given a point object in front of a plane mirror; to find by means of a drawing the position and character of the image. (a) What is the relation of the image to the object with respect to its distance from the mirror? (b) Is the image real or virtual?

45. Where must an object be placed with reference to the focus and the center of curvature of a spherical mirror in order that the image be real and (a) larger than the object? (b) smaller than the object?

46. In what two positions with reference to spherical mirrors may an object be placed so that the image is virtual? Make drawings to illustrate each.

47. The radius of curvature of a concave mirror is 30 cm. An object is placed on the principal axis 50 cm from the vertex of the mirror. Is the image (a) real or virtual? (b) erect or inverted? (c) larger or smaller than the object?

48. Under what three conditions are the images which are formed in mirrors (plane and spherical) virtual?

49. Show by means of a drawing why the virtual image in a concave spherical mirror is always larger than the object.

50. Show by means of a drawing why the virtual image in a convex spherical mirror is always smaller than the object.

51. (a) Define and explain spherical aberration in spherical mirrors. (b) Explain the formation of the "caustic."

52. Define and illustrate spherical aberration in the case of a lens, and state how it may be remedied.

53. Where must an object be placed with reference to the focus and the center of curvature of a double convex lens in order that the image be real and (a) smaller than the object? (b) same size as the object? (c) larger than the object?

54. In what position must an object be placed with reference to a convex lens in order to produce a magnified virtual image? Illustrate by a sketch.

55. Make a sketch to illustrate the position and character of the image formed by a concave lens.

56. Under what conditions does a convex lens form (a) a real image? (b) a virtual image?

57. Under what conditions does a convex lens form (a) an erect image? (b) an inverted image?

58. Under what conditions does a convex lens form (a) an image larger than the object? (b) smaller?

59. Explain by means of a diagram why the virtual image formed by a concave lens is always smaller than the object.

60. Make an outline drawing to illustrate the principle of the simple tripod microscope, and explain how it works.

61. Is the image as seen through the magnifier (Ex. 60) real or virtual?

62. What is the relation of the length of the legs of the instrument (Ex. 60) to its focal length?

63. The focal length of a simple microscope is 4 cm. What is its magnifying power?

64. Make an outline drawing to illustrate the principle of the compound microscope. Explain.

65. The tube of a given compound microscope is 6 in. in length. The focal length of the objective lens is 0.2 in. and that of the eyepiece lens 0.3 in. Find the magnifying power of this microscope.

66. Wherein does a telescope differ from a microscope?

67. At which end of the solar spectrum do we find (a) the infra-red rays? (b) the ultra-violet?

68. (a) What is a continuous spectrum? (b) a dark line spectrum? (c) bright line spectrum?

69. Under what physical conditions are each of the three kinds of spectra produced?

70. How do we know that certain metals, as sodium and iron for example, occur in the sun?

71. What is chromatic aberration, and how may it be remedied?

72. What is an achromatic lens, and in what optical instrument are such lenses used?

C

73. If the focal length of a spherical mirror is 5 in., what is the area of the sphere of which the mirror is a part?

74. A ruler 1 ft. in length is placed at right angles to, and with its middle point on the principal axis of a concave spherical mirror. The ruler is at a distance of 5 ft. from the vertex of the mirror. The radius of curvature of the mirror is 20 in. (a)

What is the focal length of the mirror? (b) How far from the mirror is the image?

75. Find the size of the image (Ex. 74).

76. An object 6 in. in height is placed on the principal axis of a convex lens at a distance of 3 ft. 4 in. from the optical center. The focal length of the lens is 20 in. How far is the image from the lens?

77. What is the size of the image (Ex. 76)?

78. Find the size of the image (Ex. 76) when the object is placed 15 in. from the lens.

79. An object 6 in. in height is placed 3 ft. from a concave lens the focal length of which is 18 in. Find (a) the position, and (b) the size of the virtual image.

80. Sometimes the wheels of an automobile, as seen on the movie screen, appear to rotate backward. How do you account for this?

81. The speed of light is 300,000 km per second, and the wave length of red light is 0.0007 mm. How many waves of light of this color will strike the eye in one second?

82. (a) Explain by means of a sketch the formation of primary and secondary rainbows. (b) Name the colors in each, in order.

83. Explain the formation of colors by interference of light waves.

84. Explain the shifting of the color bands sometimes seen in a soap film.

85. How do you account for the brilliant color effects observed when a film of oil spreads over the surface of water?

NOTE: For additional problems, see Appendix.

CHAPTER XX

ELECTRICAL RADIATIONS

Cathode Rays and X-Rays

602. Introductory. We now come to a division of Physics which not very many years ago received but little attention even in advanced text books, but which today is of enormous scientific interest to us and of much practical value as well. More investigations are probably being carried on at the present time in the field of electrical radiation and subjects relating thereto than in any other branch of physics. Not only has the whole subject of physics been profoundly affected by modern discoveries in this particular field, but the related sciences of chemistry and astronomy as well. Indeed one may say that the discovery of the electron and the properties associated with this new "ultimate electrical atom" has more profoundly affected our conception of the hidden mechanism of the physical universe than any other discovery since the days of Kepler and Newton. One has but to mention our modern notions as to the true nature of an electric current, the application of the X-ray in medical practice and in the arts, the amazing properties of radioactive substances, and the application of radium in the practical arts and sciences to indicate how wide is this new field and how significant to modern life are the subjects treated therein.

In treating the subjects of Electrical Radiations we shall in general follow the historical method, speaking in order of the Geissler tube, the Crookes tube, cathode rays, X-rays, and radioactivity.

By way of introduction, let us first consider what happens when an electrical discharge takes place in a sealed tube from which the air has been partly or almost wholly exhausted. In Fig. 598 we have represented a glass tube some two or three feet in length, and into the ends of which there have been sealed plati-

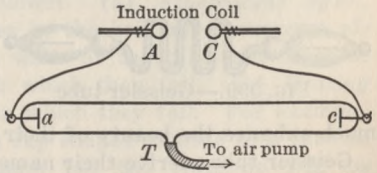


Fig. 598.—Electric discharge in an evacuated tube

num electrodes, *a* and *c*. The auxiliary tube *T* leads to a high power air pump by means of which the air may be exhausted from the glass tube in successive steps. The electrodes *a* and *c* are connected to the terminals *A* and *C* of the secondary of a relatively high potential induction coil. Assume that *A* is the positive (+) and *C* the negative (—) terminal of the coil.

Let us suppose, to begin with, that the tube is filled with air, and the induction coil is in operation. The following phenomena may be observed as the air is exhausted from the tube. (a) First, so long as the tube is filled with air no discharge between *a* and *c* occurs. Discharge, however, takes place between the knobs *A* and *C*, because the gap here is very much shorter than that between the electrodes *a c*. (b) Now start the air pump. In a few moments a discharge breaks across the gap from *a* to *c* in the tube, appearing as a ribbon of reddish light, which as the exhaustion continues, spreads out until the tube is filled with a diffused reddish light. At this stage the pressure of the air in the tube has been reduced to about 2 mm. (c) As the exhaustion continues, the color within the tube changes from red to green with intermittent dark spaces. The pressure in the tube is now about 0.5 mm. (d) And finally, as the exhaustion approaches that of a perfect vacuum the whole tube glows with a pale greenish color.

603. The Geissler Tube. *Geissler tubes* are glass tubes from which the air has been partially exhausted. Into the ends of the tubes there are sealed two short pieces of platinum wire which serve as electrodes. When these electrodes are attached to the poles of a static machine or to the terminals of an induction coil the tube becomes brilliantly lighted with colors which vary with the nature of the gas enclosed and with the kind of glass of which the tube is made. When the tube contains a trace of nitrogen the color given on discharge is violet; hydrogen, on the other hand, gives red. These tubes are often drawn into fantastic shapes, Fig. 599, which much enhance the beauty of their color effects.

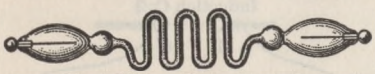


FIG. 599.—Geissler tube

Geissler tubes derive their name from Geissler, a German physicist, who invented a type of the mercury air pump, which he used in the exhaustion of these tubes. Geissler tube effects are obtained when the pressure is reduced to about 0.002 of an atmosphere, that is, 1.5 mm.

604. The Crookes Tube. The *Crookes tube*, named after Sir William Crookes, who was one of the first to work with it, is a tube from which the air has been thoroughly exhausted, the pressure being reduced to about 0.001 mm, or less. Crookes made an extensive study of the phenomena associated with electric discharge in evacuated tubes. He found, among other things, that not only does the color of the tube due to the discharge depend upon the degree of exhaustion, but also that the path of the

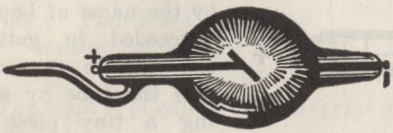


FIG. 600.—Crookes tube



Crookes

discharge from the cathode (negative electrode) depends upon the degree to which the air has been removed.

Crookes tubes usually possess a bulb-like form, as shown in Fig. 600.

605. Cathode Rays. *Cathode rays* are streams of negatively charged particles (electrons) which are shot off from the cathode in a highly exhausted tube upon the application of a high potential difference to the electrodes of the tube. These rays manifest themselves in various ways as follows: (a) *Fluorescent effect.* Cathode rays manifest themselves by the brilliant fluorescent effects which they produce in the walls of the tube and in other substances within the tube upon which they fall. (b) *Heating effect.* The rays heat bodies upon which they fall. For example, if the stream of electrons from a cup shaped cathode in a tube be focused upon a piece of platinum foil the platinum will be heated white hot. (c) *Mechanical effect.* In the tube shown in Fig. 601 there is a light wheel having mica vanes and so adjusted as to roll on smooth glass rails. When a stream of electrons is directed from the cathode so as to strike the upper vanes

of the wheel it will rotate, rolling along the glass rails. When the current is reversed the direction of rotation of the wheel is reversed. (d) *Magnetic effect.* Cathode rays are deflected from their course by a magnet, as shown in Fig. 602. The stream of electrons constitute an electric current, and this current is surrounded by a magnetic field, hence its deflection by a magnet.

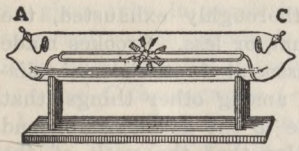


FIG. 601.—Mechanical effect of cathode rays

Following the investigations of Crookes, a Hungarian physicist by the name of Lenard succeeded in getting the cathode rays outside the tube by sealing a tiny piece of thin aluminum over a hole in one end. These cathode rays which passed out of the tube through the aluminum window are called *Lenard rays.*

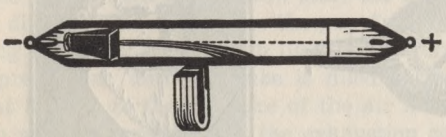


FIG. 602.—Cathode rays deflected by a magnet

606. X-Rays. X-rays, or Röntgen rays as they are sometimes called, were discovered by Wilhelm Konrad Röntgen, a German physicist, in 1895. When cathode rays (streams of electrons) fall upon an object, as for example the metal reflector of a Crookes tube, Fig. 603, they give rise to X-rays, somewhat analogous to the manner in which a stone dropped into water gives rise to water waves. It is important to note that cathode rays are quite different from X-rays. The cathode ray is no more the X-ray than the stone is the water wave; one is the cause of the other.



Röntgen

X-rays do not carry electrical charges, nor are they deflected by a magnet. X-rays consist of extremely short waves similar to but very much shorter than light waves. The wave length of ordinary yellow light is 0.0000589 cm; the wave length of the shortest X-rays is about one hundred millionth of a centimeter,

that is, 0.00000001 cm, representing a vibration rate of 3,000,000,000,000,000,000 per second.

When an electron, in a Crooke's tube for example, strikes a target, X-rays are set up as a result of the impact. Now there is always a constant ratio between that portion of the kinetic energy of the electron which is converted into the energy of the X-rays and the frequency of the rays produced. This relation may be expressed in terms of Planck's equation as

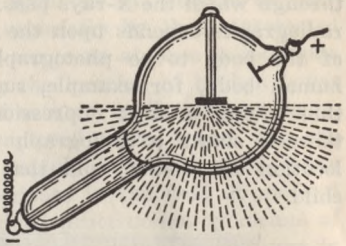


FIG. 603.—X-ray tube

amount of energy (quantum) lost by the electron = $h \times$ *frequency of resulting X-rays*

The term h is called the *universal constant*, and is equal to 6.56×10^{-27}.

607. Properties of X-rays. Röntgen rays possess certain remarkable properties which may be stated briefly as follows: (a) Röntgen rays excite phosphorescence in a large number of substances. (b) Gases through which Röntgen rays pass are ionized and become conductors of electricity. Thus if a charged electroscope be placed in the neighborhood of an active X-ray tube it will be observed that the gold leaves collapse, the charge of the instrument being rapidly carried away by the ionized air. (c) These rays have great penetrating power, being able to pass through bodies of considerable thickness. Different substances absorb the rays in different degrees, as is well illustrated by the parts of the human body. The bones, for example, absorb the rays more strongly than do the fleshy parts. Metals also absorb the X-rays more strongly than do the non-metals, such as wood or paper.

The penetrating power of Röntgen rays is determined largely by the pressure of the gas within the tube. A high degree of exhaustion, and consequent low pressure of the gas in the tube, gives rays of high penetrating power, known as "hard rays"; somewhat lower exhaustion gives rays of less penetrating power, which are known as "soft rays."

Röntgen rays produce photographic action somewhat similar to that due to light, giving rise to the X-ray photograph or *radiograph.*

608. The Radiograph. The so-called X-ray photograph is really a shadow picture cast upon the photographic plate by the body through which the X-rays pass. The possibility of obtaining these radiographs depends upon the absorbing power of different parts of the body to be photographed. The denser portions of the human body, for example, such as the bones, absorb the rays, thus giving a dark impression upon the picture. In Fig. 604 there is shown a radiograph which revealed to the surgeon the location of a safety pin that was accidentally swallowed by a child.

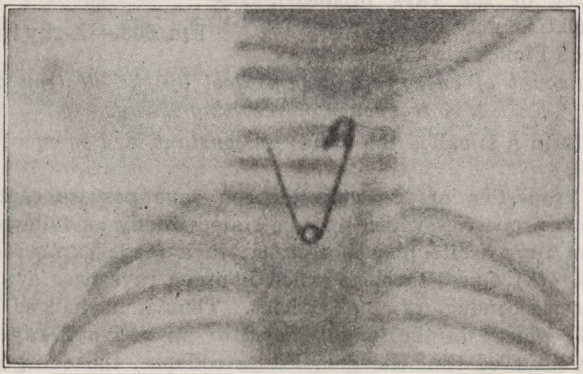


FIG. 604.—Radiograph revealing to the surgeon the location of a safety pin which was accidentally swallowed by a child

609. The Fluoroscope. A *fluoroscope* is an apparatus which enables us to study the effects of X-rays without the use of photographic plates. It consists of a box as shown in Fig. 605, the small end of which is so constructed as to fit closely around the eyes. Over the large end is placed a fluorescent screen, made usually of platino-barium-cyanide If an object, such as the hand, be placed between this screen and an X-ray tube the outlines of the denser portions of the object may be seen distinctly.

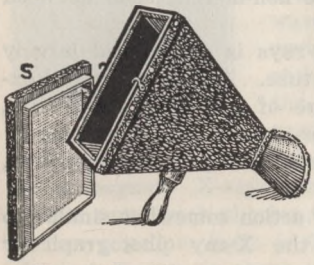


FIG. 605.—Fluoroscope

610. What X-Rays Show. The discovery of the X-ray was announced by Röntgen in 1895. The possibilities of the application of this new and wonderful agent in the domain of photography were at once recognized especially by those interested in medical science. Physicians and surgeons at once seized upon the X-ray as an aid in determining the location of broken bones and the extent and character of fractures in the human body. Naturally at first the methods employed were crude, but the technique of the X-ray work has rapidly improved until today rontgenology is a well established science in the field of medical diagnostic practice. Today the medical practitioner by means of the radiograph is not only able to locate a break or fracture in a bone or the position of dense foreign substances in the body, but also he is able by means of the radiograph to study diseased conditions of the tissues such as abnormal growths within the body.

The dentist also has found that the X-ray is practically the only means whereby he can be sure that he is properly diagnosing certain ailments of his patient. Defective teeth are sometimes due to abscesses and other growths forming at the roots, which are not visible in any other way by physical examination, except by the X-ray. Such abscesses are the cause of serious organic and systemic disorders and ailments, such as rheumatism, neuralgia, stomach and intestinal troubles, and many other ills which would be difficult to determine definitely and to treat correctly without the helpful aid of the radiograph.

The X-ray is no longer confined to hospital and physicians' diagnostic requirements, but it has been used in the past few years, especially during the world war, for many hitherto undreamt of purposes. In fact, one could write a book on the industrial applications of X-rays which would read like a romance. For example, during the war and after, great use was made of the X-ray in making detailed and critical studies of the physical condition of the materials which entered into the construction of airplanes. A single weak place in a bar or in the engine of an airplane may mean the death of an aviator, and it is, therefore, necessary to examine critically every detail of the construction of aircraft.

Undoubtedly many railway accidents are caused by defects in rails and in the car wheels, defects which are invisible to the unaided eye. It frequently happens that car wheels which on ordinary inspection appear to be all right, and which withstand

possibly a fair mechanical test, inherently possess a weakness, perhaps in the form of air pockets or invisible cracks caused by strains in cooling when they were cast, defects which only the X-ray can and does detect at once in a minimum of time.

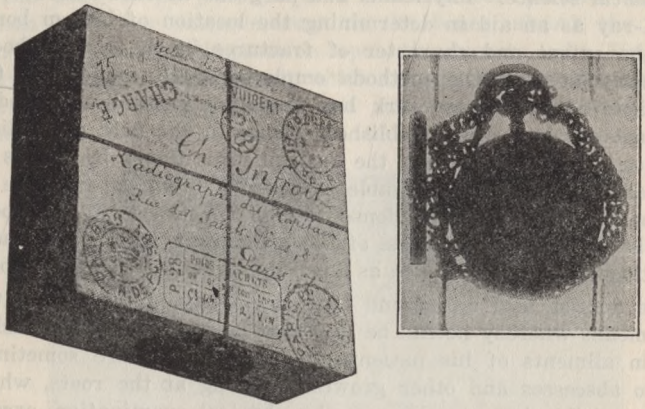


FIG. 606.—Sealed package and contents, as revealed by the fluoroscope

Custom house officers often use the X-ray to examine the contents of imported packages, Fig. 606, showing such a package and what it contained, as revealed by the fluoroscope.

More romantic perhaps is the application of the X-ray for discovering the presence of pearls or the nuclei of pearls starting to grow in oysters. Previously many thousands of oysters were destroyed annually by being opened in an effort to determine the presence of pearls or pearl nuclei. All this is now obviated, thanks to the X-ray. If the oyster does not show up with a pearl nucleus it is replanted in the water. Here it develops to maturity. If an undeveloped pearl is detected by the X-ray pearl expert, then the oyster is at once laid aside for further treatment.

611. The Coolidge Tube. Many remarkable X-ray and cathode ray effects have been produced by means of a new vacuum tube which was invented by Dr. W. D. Coolidge, the discoverer of a way to make strong and ductile filaments of tungsten (Art. 408). We learned in Art. 605 that when a high voltage electrical discharge is passed through a tube into which a metallic terminal has been sealed at each end, and from which the air has been

almost completely exhausted, a rapidly moving stream of electrons results. These streams are called cathode rays, and when they hit a piece of some dense metal, such as platinum or tungsten, X-rays are given off.

Instead of using an ordinary piece of metal for the cathode, the Coolidge X-ray tube makes use of a tiny electric-light filament. So complete is the vacuum in the Coolidge tube that with the highest voltages no current would pass through it when this little filament is cold. But when the filament is heated by a low-voltage current, no more than that used in an automobile headlight, electrons are given off, and when a high voltage is applied to the tube these electrons are driven against the heavy metal target and we have a steady and reliable source of X-rays.

And not only does the Coolidge tube furnish a reliable and effective means of producing X-rays but when it is desired, Lenard rays (streams of electrons outside the tube) may be produced in abundance. While Lenard was confined to a thin aluminum window an eighth of an inch in diameter, the new tube uses as a window a sheet of nickel five ten-thousandths of an inch thick and three inches in diameter. Through this window come quantities of cathode rays far in excess of any that Crookes obtained inside his vacuum tube.

RADIOACTIVITY

612. Introductory. Nearly everyone nowadays has heard of *radium* and the remarkable properties which it possesses. The story of the discovery of this wonderful element is one of the most interesting in the annals of modern science. In 1896 Henri Becquerel, of Paris, accidentally discovered that the mineral uranium possessed the power of affecting a photographic plate. Following this discovery he wrapped a photographic plate in a piece of perfectly black paper, laid a coin on top of the paper, and suspended above the coin a small quantity of uranium. He then set the whole away in a dark room for several days. When he developed the photographic plate he found upon it a shadow picture of the coin similar to that produced by the Röntgen rays from an



Becquerel

X-ray tube. He concluded, therefore, that uranium possesses the property of emitting rays of some sort which have the power of penetrating opaque objects and of affecting photographic plates, just as X-rays do. He found also, that these rays are like X-rays in that they discharge electrically charged bodies upon which they fall. Becquerel's discovery created a profound sensation among physicists and chemists, and at once stimulated a search for other substances besides uranium which might possess the property of emitting these new rays, which soon became known as Becquerel rays. A number of other such substances were found.

Any substance which spontaneously emits radiations similar to those which come from uranium is called radioactive, and it is said to possess the property of *radioactivity*.

613. Becquerel Rays. A study of radioactive substances revealed the fact that they not only emit rays which are capable of affecting a photographic plate, but they also project from themselves electrically charged particles. In 1899 Professor Rutherford, then of McGill University, Canada, found that the Becquerel rays consist of three different kinds of radiations. These three types of radiations he named alpha, beta, and gamma rays, taking these names from the first three letters of the Greek alphabet, namely, α (alpha), β (beta), and γ (gamma).

When Becquerel rays were submitted to a strong magnetic field it was found that the gamma rays were unaffected, while the alpha rays were deflected in one direction and the beta rays in the opposite direction, Fig. 607.

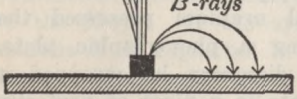


FIG. 607.—Alpha, beta and gamma rays

The *alpha* rays possess relatively small penetrating power. They consist of positively charged particles having a mass which is calculated to be about 4 times that of the hydrogen atom. Rutherford concluded that the particles of the alpha rays are positively charged atoms of *helium*.

The *beta* rays were found to be identical in all respects, with cathode rays; that is, the negatively charged particles of the beta rays are electrons.

The *gamma* rays, which are unaffected by a magnetic field, consist of very short waves similar to those of the X-rays. In

other words, they are very short X-rays. These rays are very penetrating and produce photographic effects.

614. Radium. Shortly after the discovery of the Becquerel rays Madam Curie, who was then working with her husband in one of the physical laboratories of Paris, began an investigation of all the known elements, other than uranium, to find whether any of them possessed the remarkable property of radioactivity. She found that the element thorium, which is one of the chief constituents of Welsbach gas mantles, is capable, together with its compounds, of producing the same effect as uranium. In connection with this investigation Madame Curie noticed that pitchblende, the crude ore from which uranium is extracted, and which consists largely of uranium oxide, would discharge her electroscope about four times as fast as pure uranium. She inferred that the radioactivity of pitchblende could not be due solely to the uranium contained in it, and that this ore must therefore contain some unknown element which has the property of emitting Becquerel rays more powerfully than uranium or thorium.



Madame Curie

After prolonged experimenting she obtained from several tons of pitchblende a few milligrams of a substance more than a million times as active as thorium or uranium. She called this new substance *radium*.

As has just been stated, radium is more than a million times as radioactive as uranium, emitting alpha, beta, and gamma rays. When a radium tube is brought near a charged electroscope the leaves at once collapse. This is because the gamma rays emitted by the radium ionize the air in the electroscope causing it to become electrically conductive. Furthermore, radium is continually decomposing into other substances, the decomposition being accompanied by the evolution of heat.

615. Radium as a Medical Agent. Radium is obtained as a chemical compound, usually as radium chloride or radium bromide. In this country, radium is obtained mainly from Colorado "carnotite" ore. From a whole ton of high grade Colorado carnotite ore we get only from five to seven milligrams of radium element.

Locked up in radium is one of the greatest forces that the world knows. Chemically it is very active, even being so powerful as to destroy living tissue and bone when brought for any length of time in contact with the body. An ounce of it carried in the vest pocket in an ordinary glass tube, would kill a man in ten hours by destroying his tissues and bones. If a man carried one gram of it in his pocket two days, serious results would likewise follow. The careless handling of glass tubes containing only twenty-five milligrams of radium will cause deep wounds on the fingers of an operator. Notwithstanding its great chemical activity, however, radium has been successfully employed in medical practice in the treatment of certain diseases. Radium if carelessly handled will destroy tissues and bones, yet by the proper application under the direction of a medical expert it can be used so as to do no harm to healthy tissue, while destroying certain unhealthy tissues or malignant growths. It has been used to remove birthmarks and to reduce the scars inflicted on soldiers in the late war. Undoubtedly in the future we shall learn much more of its value as an aid to the physician and the surgeon.

616. Radium as an Illuminating Agent. Radium offers interesting and attractive features as a commercial illuminating agent, and undoubtedly in the future, when the production of radium salt from the natural ores has been developed on a much larger scale than at present radium illuminating paints will be used extensively.

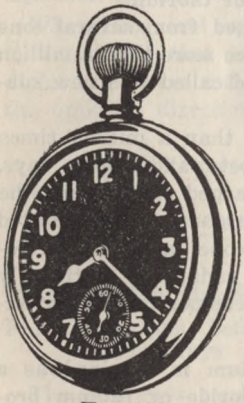


FIG. 608
"Radiolite" watch

One of the most familiar examples of radium illumination is that furnished by the modern "radiolite" watch, the hands and figures on the dials of which are so illuminated as to make it possible to tell the time in the dark, Fig. 608. When only a few millionths of a gram of radium is mixed with zinc sulphide, varnish, and certain other adhesive substances, it gives enough material to illuminate forty or fifty watches.

The commercial world was quick to see the many valuable uses that can be made of material which will glow in the dark. It was

used on the dials of airplanes which flew over enemy camps at night, when the exposure of light of another character would have made the airplane a target for shells.

Today, the dashboard instruments of automobiles, such as the speedometer, clock, and various gauges, tell the driver the condition of his car without the use of the electric light. Radium is practically the only means of illuminating motorcycle speedometers at night, since the motorcycle does not carry instrument lights. The steersman at the wheel of ocean going vessels can now be independent of artificial lights at night. Radium illumination is particularly good for these purposes, because when the eye has consulted a dial which is brightly lighted with electricity the vision is affected for a considerable period, so that the eye cannot look off into the dark with a normal degree of certainty.

Steam gauges, water-level glasses, and air-pressure gauges on locomotives and stationary engines are nowadays lighted with radium; so also is the pressure-dial of the heater in the dark or semi-dark cellar of the modern dwelling. Radium lighted "Danger" signs and indicating arrows have been made for use in mines to protect miners from the danger of explosion of coal gas. Radium paint is also used for illuminating door plates.

RECENT INVESTIGATIONS AS TO ATOMIC STRUCTURE AND RADIATIONS

617. Atomic Structure. In our study of atoms thus far, we have learned that an atom consists of a positively charged nucleus around which there are one or more free electrons. Neils Bohr, a Danish physicist, advanced the theory that the electrons rotate around the nucleus somewhat as the planets revolve around the sun. While this theory has been, and is being modified as researches are carried forward, the Bohr atom, with its positive nucleus and associated free electrons, gives a fairly good picture of what an atom is like. We know that there are electrons spaced around the nucleus. The important question is: What is the nucleus made of?

The nucleus of an atom may consist of one proton, as in the case of hydrogen, or it may consist of a number of protons associated with which there are a *lesser* number of bound electrons,

and there may also be other elementary units. The following facts are known about electrons and protons.

(a) Every electron has a negative charge, which is always the same.

(b) Every proton has a positive charge, which is equal in magnitude to the negative charge of the electrons.

(c) The mass of a proton is about 1850 times as great as that of an electron; therefore, the mass of an atom is dependent mainly upon the number of protons in the nucleus.

(d) Electrons are free to move. An electric current consists of a stream of moving electrons.

(e) Protons in solid conductors are not free to move. The only way a proton can move is for the atom to move, as in electrolysis, or when the nucleus explodes, as in radioactivity, or when the nucleus is knocked to pieces by laboratory methods. Under ordinary circumstances a proton does not leave the nucleus.

As has just been stated the nucleus of an atom may consist of protons and electrons closely packed together, the number of protons always being greater than the number of associated electrons.

The number of positive charges which a nucleus carries is equal to the number of protons in it *minus* the number of associated electrons. This is shown in the following table:

Atom	*Nucleus*	*Positive charges*	*Satellites*
Hydrogen	1 proton	1	1 electron
Helium	4 protons, 2 electrons	2	2 electrons
Beryllium	9 protons, 6 electrons	3	3 electrons
Carbon	12 protons, 6 electrons	6	6 electrons

618. Disintegration of the Atom. Radioactive matter is in the process of disintegrating, that is, of breaking down. When energy is liberated, as in the case of radioactivity, the heavy atom breaks down into simpler and lighter atoms. For example, when Uranium, which has an atomic weight of 238, disintegrates, giving off alpha rays, Ionium (atomic weight, 230) is formed; and so from Ionium there comes Radium (atomic weight, 226); and from Radium through a series of disintegrations, comes Lead (atomic weight, 207). Lead appears to be the end product of this series of disintegrations.

Radioactivity seems to be nature's way of breaking up atoms. In order to find out what atoms are made of, and especially in order to investigate the structure and composition of the nucleus,

scientists have devised various methods of knocking atoms to pieces. One method of doing this is to shoot the atom to pieces, so to speak, by firing electrically charged bullets into it. The electrical bullet used is a positively charged proton which is fired off with enormous velocity from an electrode having a potential of several million volts.

The nucleus of a hydrogen atom consists of a single proton. In the case of other elements the nucleus is made up of protons and electrons closely associated. When such a nucleus is broken up, there may be released other elementary units which are known as positrons and neutrons. A *positron* appears to be of the nature of an electron with a positive charge, and a *neutron* is of the nature of a proton with no charge at all. Investigations in this particular field, however, are so recent that one is not justified at present in making statements as to the exact nature of these new units.

619. Atomic Energy. When radioactive elements disintegrate energy is liberated. A gram of radium gives off about 120 calories of heat energy per hour. J. J. Thomson has estimated that if all the energy contained in the atoms in one gram of hydrogen could be liberated more than 600 million foot pounds of energy would be given off. It is evident then that an enormous amount of energy is contained in atoms; in fact, one of the greatest storehouses of energy in the world is that which is locked up within atoms.

The atomic energy of radium is used in the treatment of cancer and for other purposes. The energy used for such purposes is only an infinitesimal part of the total atomic energy. Will man ever be able to use atomic energy for mechanical purposes? As yet no one knows the answer to this question. It is doubtful that it will be possible because of the difficulty of liberating this energy. However, some new way of unlocking this atomic energy may be found. Science has accomplished wonders in the past and it will, no doubt, continue to do so.

620. Cosmic Rays. Cosmic rays are sometimes called Millikan rays in honor of their discoverer, Professor Robert Millikan. These rays come to the earth in all directions from some outside source or sources. Since at a given place on the earth there is apparently no variation in their intensity between day and night, one would infer that their source is not the sun or any definite

group of stars. They appear to come "from the depths of space."

The discovery of cosmic rays has raised a number of important scientific questions. The first concerns the nature of these rays, that is: Are cosmic rays waves or are they particles? Professor Millikan has advanced the theory that cosmic rays are waves while Professor Arthur Compton has suggested that they are of the nature of particles. The second question relates to the origin of the cosmic waves. A third important question raises an inquiry as to what can be learned from cosmic rays as to the relation between mass and energy.

The penetrating power of cosmic waves is very much greater than that of x-rays, and if cosmic rays are to be considered as waves, the wave lengths must be much shorter than those of light waves or of x-rays, as is shown in the following table.

Kind of Wave	*Wave Length in cm*
Visible spectrum, from	0.000,074
to	0.000,040
X-rays, from	0.000,000,12
to	0.000,000,001
Cosmic rays, from	0.000,000,000,005
to	0.000,000,000,003

The wave lengths of cosmic rays, as given above, are computed from their penetrating power, using the quantum relation.

621. Mass and Energy. As has been suggested, a further study of cosmic rays may solve or help to solve the problem of the relation between mass and energy. In the beginning of our study of physics, we learned that mass is the quantity of matter in a body and matter was provisionally defined as anything that occupies space and has weight. This definition is all right for all ordinary purposes. When, however, we come to consider such exceedingly small particles as electrons and such high frequencies as those of cosmic rays, this definition is far from satisfactory. For example, it has been found that the mass of an electron appears to increase as its velocity increases. In other words, the quantity of matter in this elementary particle is not constant. Also, we have every reason to believe that chemical elements are built up mainly of two things, electrons and protons. In this building process, however, there appears in some cases to be a loss of mass. If this is true, then the quantity of matter in the universe is not constant as was once believed.

These facts, together with those presented in the study of radioactivity, lead physicists to conclude that, so far as the present status of our knowledge is concerned, when an atom loses energy it loses mass, and when it gains energy it gains mass. According to Einstein, *mass is a form of energy*. That is, when there is a change in the internal energy of an atom there must be a change of mass.

These conclusions are of vital interest to scientists. They do not, however, affect the practical applications of our mass and energy concepts as used in everyday life.

SUMMARY

1. Cathode rays are streams of electrons which are projected from the cathode of a highly exhausted tube. A cathode ray is of the nature of an electric current, and as such possesses heating, lighting, and magnetic properties.

2. Lenard rays are cathode rays which have passed out of the vacuum tube through a thin metal window.

3. When cathode rays strike upon certain dense substances, such as platinum and tungsten, they give rise to X-rays.

4. X-rays are extremely short ether waves of the same nature as light waves, only very much shorter.

5. X-rays that have high penetrating power are said to be "hard"; those of low penetrating power are "soft."

6. The Coolidge X-ray tube is a vacuum tube of an extremely high degree of exhaustion, in which the cathode is a tiny tungsten filament, which when hot gives off electrons freely. The characteristic feature of this tube is the use of a hot filament as the cathode.

7. A radioactive substance is one that spontaneously gives off Becquerel rays.

8. Becquerel rays consist of three kinds of radiations: alpha rays, beta rays, and gamma rays.

9. Alpha rays are composed of positively charged atoms of helium.

10. Beta rays are composed of negatively charged electrons.

11. Gamma rays are very short X-rays.

12. Radium is the most radioactive substance known; it is an element which is in a continuous state of decomposition, breaking down into other elements, with the evaluation of heat.

REVIEW EXERCISES

1. Wherein does a Crookes tube differ from a Geissler tube?

2. What are cathode rays? Name the principal effects of cathode rays.

3. How do you account for the fact that cathode rays are deflected by a magnet?

4. What are Lenard rays, and wherein do they differ from cathode rays?

5. What is an X-ray, and how is it produced? Wherein do X-rays differ from cathode rays?

6. What are "hard" X-rays, and how are they produced? What are "soft" rays?

7. What is a radiograph? Why do the bones of the body show up so much more clearly in a radiograph than the muscles?

8. What use do physicians and dentists make of X-rays?

9. What are some of the commercial and industrial uses of X-rays?

10. Wherein does a Coolidge tube differ from a Crookes tube?

11. Why is an electroscope discharged when it is brought in the neighborhood of X-rays?

12. Give an account of Becquerel's discovery. What is a radioactive substance? Of what are Becquerel rays composed?

13. What is an alpha ray? a beta ray? a gamma ray?

14. Are alpha rays attracted or repelled by a positively charged body? Why?

15. How are beta rays affected by a positively charged body? Why?

16. Give an account of the discovery of radium.

17. Name some of the main properties of radium.

18. How do you account for the fact that when a tube containing some radium salt is brought near a charged electroscope the leaves at once collapse?

APPENDIX

TABLES

TABLE I, UNITS OF LENGTH

Metric Units

Fractions		*Multiples*	
1 decimeter (dm)	= .1 meter	1 dekameter (dkm)	= 10 meters
1 centimeter (cm)	= .01 meter	1 hectometer (hm)	= 100 meters
1 milimeter (mm)	= .001 meter	1 kilometer (km)	= 1000 meters

English Units

12 inches	= 1 foot	1000 mils	= 1 inch
3 feet	= 1 yard	4 inches	= 1 hand
16.5 feet	= 1 rod	9 inches	= 1 span
320 rods	= 1 statute mile	6 feet	= 1 fathom
1 statute mile	= 5280 feet		
1 nautical mile	= 6080 feet		

Conversion Factors

Metric to English	*English to Metric*
1 kilometer = 0.62137 mi.	1 mile = 1.60935 km
1 meter = 3.28083 ft.	1 foot = 0.3048 m
1 centimeter = 0.3937 in.	1 inch = 2.54 cm

TABLE II, AREA AND VOLUME

English Units

144 square inches	= 1 square foot	36 square miles	= 1 township
9 square feet	= 1 square yard	1728 cubic inches	= 1 cubic foot
30¼ square yards	= 1 square rod	27 cubic feet	= 1 cubic yard
160 square rods	= 1 acre	128 cubic feet	= 1 cord
640 acres	= 1 square mile		

TABLE III, CAPACITY

English Units

Dry Measure		*Liquid Measure*	
2 pints	= 1 quart	4 gills	= 1 pint
8 quarts	= 1 peck	2 pints	= 1 quart
4 pecks	= 1 bushel	4 quarts	= 1 gallon
1 bushel	= 2150.42 cu. in.	31½ gallons	= 1 barrel
1 standard barrel	= 7056 cu. in.	1 gallon	= 231 cu. in.

Conversion Factors

Metric to English	*English to Metric*
1 liter = 1.057 qts., liquid measure	1 quart, L.M. = 946.3 cc
1 liter = 0.908 qts., dry measure	1 quart, D.M. = 1001.2 cc

TABLE IV, MASS

Metric Units		*English Units, Avoirdupois*	
1 gram (g)	= .001 kilogram (kgm)	437.5 grains	= 1 ounce
1 decigram (dg)	= .1 gram	16 ounces	= 1 pound
1 centigram (cg)	= .01 gram	1 pound	= 7000 grains
1 milligram (mg)	= .001 gram	2000 pounds	= 1 ton

Conversion Factors

Metric to English	*English to Metric*
1 kilogram = 2.2046 lb.	1 pound = 0.4536 kg
1 gram = 15.432 grains	1 grain = 0.0648 g

U. S. Coins

5 cent piece (nickel)	= 5.0 grams
10 cent piece (dime)	= 2.5 grams
25 cent piece (quarter)	= 6.25 grams
50 cent piece (half dollar)	= 12.5 grams

Karat and Carat

1 karat (fineness of gold) = 1/24 by weight of pure gold
1 carat (mass of precious stones) = 200 milligrams

TABLE V, WIRE GAUGE, B. & S.

Gauge No.	*Diam. in mm*	*Diam. in mils*	*Gauge No.*	*Diam. in mm*	*Diam. in mils*
0000	11.684	460.00	19	.899	35.39
000	10.405	409.64	20	.812	31.96
00	9.266	364.80	21	.723	28.46
0	8.254	324.95	22	.644	25.35
1	7.348	289.30	23	.573	22.57
2	6.544	257.63	24	.511	20.10
3	5.827	229.42	25	.455	17.90
4	5.189	204.31	26	.405	15.94
5	4.621	181.94	27	.361	14.19
6	4.115	162.02	28	.321	12.64
7	3.665	144.28	29	.286	11.26
8	3.264	128.49	30	.255	10.03
9	2.907	114.43	31	.227	8.93
10	2.588	101.89	32	.202	7.95
11	2.305	90.74	33	.180	7.08
12	2.053	80.81	34	.160	6.30
13	1.828	71.96	35	.143	5.61
14	1.628	64.01	36	.127	5.00
15	1.450	57.07	37	.113	4.45
16	1.291	50.82	38	.101	3.96
17	1.150	45.26	39	.090	3.53
18	1.024	40.30	40	.080	3.14

TABLE VI, EQUIVALENTS AND EQUATIONS

$\pi = 3.1416$; $\pi^2 = 9.8696 = (9.87)$; $\sqrt{\pi} = 1.7724$; $\log \pi = 0.49715$.

Circle: *circumference* $= 2\pi r$; *area* $= \pi r^2$.

Cylinder: *lateral area* $= 2\pi rh$; *volume* $= \pi r^2 h$.

Cone: *lateral area* $= 1/2$ (circumference $\times$ slant height); *volume* $= 1/3 \times$ (area base $\times$ height).

Sphere: *area* $= 4\pi r^2$; *volume* $= 4/3\pi r^3$.

Pressure of one atmosphere = 76 cm mercury = 30 in. mercury = 1033.3 g per cm^2 = 1,012,634 dynes per cm^2 = 14.7 lb. per sq. in.

Mass of 1 cc of pure water at 4° C = 1 gram.

Mass of 1 cu. ft. of pure water at 4° C = 62.3565 lb. = (62.4 lb.).

Mass of 1 gallon of pure water at 4° C = 8.33585 lb. = (8.3 lb.).

TABLE VII, DENSITIES IN GRAMS PER CM³

Air, at 0° C and 76 cm pressure	0.00129	Iron, wrought	7.8
Alcohol	0.80	Ivory	1.82
Aluminum	2.67	Kerosene	0.8
Antimony	6.72	Lead	11.30
Beeswax	0.96	Magnesium	1.75
Bismuth	9.82	Marble	2.72
Brass	8.5	Mercury, at 0° C	13.596
Charcoal	1.60	Milk	1.03
Coal	1.3 to 1.80	Nickel	8.9
Copper	8.9	Olive oil	0.92
Cork	0.24	Osmium	22.5
Diamond	3.53	Paraffin	0.90
Ether	0.74	Platinum	21.50
Gasoline	0.75	Silver	10.56
German silver	8.43	Steel	7.82
Glass, crown	2.60	Sulphuric acid	1.84
Glass, flint	3.70	Sulphur	2.03
Glycerine	1.26	Sugar	1.59
Gold	19.30	Tin	7.29
Granite	2.70	Tungsten	19.60
Human body	0.89	Water, at 0° C	0.999
Ice	0.92	Water, at 4° C	1.00
Iron, gray cast	7.10	Water, sea	1.03
		Zinc	7.00

TABLE VIII

Surface Tension in Dynes per Cm

Mercury	547
Water	74
Olive oil	32
Turpentine	27
Alcohol	23

TABLE IX

Values of G in Cm per Second per Second

Boston, Mass.	980.38	Washington, D. C.	980.10
Ithaca, N. Y.	980.29	Cincinnati, O.	979.99
Chicago, Ill.	980.26	Charlotteville, Va.	979.92
Cleveland, O.	980.23	Denver, Col.	979.60
Philadelphia, Pa.	980.18	Pike's Peak, Col.	978.94

TABLE X

Specific Heats

Air	0.237	Iron	0.113
Alcohol	0.602	Lead	0.03
Aluminum	0.22	Marble	0.21
Brass	0.094	Mercury	0.033
Copper	0.09	Silver	0.056
Glycerine	0.55	Steam, 100° C, 76 cm.	0.48
Glass	0.2	Water	1.00
Ice	0.5	Zinc	0.094

TABLE XI

Coefficients of Linear Expansion in Cm per Degree per Cm

Aluminum	0.000023	Iron and steel	0.000012
Brass	0.000018	Lead	0.000028
Copper	0.000017	Platinum	0.000003
Glass	0.000008	Silver	0.000019
Gold	0.000014	Zinc	0.000029

TABLE XII

Melting Points

Aluminum	657° C	Lead	327° C
Beeswax	62	Mercury	—38.8
Butter	33	Paraffin	45 to 50
Copper	1084	Platinum	1778
Glass	1000 to 1400	Rose's fusible metal	94
Gold	1063	Solder, soft	225
Ice	0	Sulphur	115
Iridium	2200	Tin	232
Iron	1100 to 1200	Tungsten	2950

TABLE XIII

Boiling Points

Alcohol	78° C	Glycerine	290° C
Ammonia	38	Mercury	357
Benzene	80	Toluene	110
Carbon disulphide	46	Turpentine	160
Chloroform	61	Water	100

TABLE XIV

NUMBER OF GRAMS OF WATER VAPOR REQUIRED TO SATURATE THE AIR, PER CUBIC METER

0° C.....4.835 g	10° C..... 9.330 g	20° C.....17.118 g
1° 5.176	11° 9.935	21° 18.143
2° 5.538	12° 10.574	22° 19.222
3° 5.922	13° 11.249	23° 20.355
4° 6.330	14° 11.961	24° 21.546
5° 6.761	15° 12.712	25° 22.796
6° 7.219	16° 13.505	26° 24.109
7° 7.703	17° 14.339	27° 25.487
8° 8.216	18° 15.218	28° 26.933
9° 8.757	19° 16.144	29° 28.450

TABLE XV

THERMAL CONDUCTIVITIES IN CALORIES PER CM^2 PER SECOND PER DEGREE CENTIGRADE PER CM

Metals

Aluminum 0.5	Lead 0.08
Brass 0.26	Mercury 0.02
Copper 0.92	Silver 1.00
Iron, cast.....0.11	Zinc 0.27

Other Substances

Asbestos 0.0006	Glass 0.00250
Cotton, wool.....0.00004	Sawdust 0.00012
Cotton cloth.....0.00055	Silk 0.00022
Cork 0.00013	Wood, dry.....0.00040
Felt 0.00009	
Flannel 0.00023	

TABLE XVI

COMPARISON OF HEAT VALUES OF FUELS

	B.t.u. per lb.		*B.t.u. per lb.*
Wood, soft	7,000	Fuel oil	18,000
Wood, hard	7,000	Kerosene	18,000
Coal, bituminous	14,000	Gasoline	19,000
Coal, semi-bituminous..	15,000	Alcohol, wood.....	9,000
Coal, hard	13,000		*B.t.u. per cu. ft.*
Charcoal	16,000		
Coke	12,000	Natural gas.....	1,000
Alcohol, ethyl.....	12,000	Manufactured gas.....	600

TABLE XVII

Resistivity Values at 0° C

Conductor	*k in ohms per cm*	*k′ in ohms per mil-foot*	*Conductor*	*k in ohms per cm*	*k′ in ohms per mil-foot*
Aluminum ..	2.6×10^{-6}	17.5	Silver	1.5×10^{-6}	9.5
Copper	1.6×10^{-6}	10.1	Manganin .	42×10^{-6}	260.0
Gold	2.2×10^{-6}	12.6	Mercury ..	94×10^{-6}	566.0
Iron	9.7×10^{-6}	58.3	Nickel	12×10^{-6}	75.0
German silver	26.6×10^{-6}	125.7	Platinum ..	10×10^{-6}	60.0

TABLE XVIII

Coefficients of Sound Absorption

Taking the absorbing power of 1 square foot of open window as unity, the coefficients of the following substances may be expressed in absorbing power per square foot:

Substance	*Coefficient*	*Substance*	*Coefficient*
Open window	1.00	Wood, plain.....	0.061
Hair felt, 1 in. thick..	.55	Plaster on lath.....	0.033
Artificial stone	.36	Cork floor tile.....	0.030
Celotex, half in. thick..	.31	Wood, varnished	0.030
Insulite, " " " ..	.31	Linoleum	0.030
Stage opening	.25	Window glass.....	0.027
Curtains	.23	Plaster on tile.....	0.025
Carpets	.20	Brick wall	0.025
Cloth, cretonne.....	.19	Marble	0.010

TABLE XIX

Indices of Refraction

Ice	1.31	Benzene	1.50
Water	1.33	Crown glass	1.52
Alcohol	1.36	Flint glass	1.62
Turpentine	1.47	Carbon disulphide	1.64
Canada balsam	1.50	Diamond	2.47

TABLE XX

WAVE LENGTH AND FREQUENCY OF THE COLORS OF THE SOLAR SPECTRUM

Color		*Wave length in mm*	*No. vibrations per second*
Red	A	.000760	395,000,000,000,000
Orange	C	.000656	458,000,000,000,000
Yellow	D	.000589	510,000,000,000,000
Green	E	.000527	570,000,000,000,000
Blue	F	.000486	618,000,000,000,000
Indigo	G	.000431	697,000,000,000,000
Violet	H	.000397	760,000,000,000,000

SUPPLEMENTARY PROBLEMS

CHAPTER I. **1.** The train known as the "Twentieth Century Limited" of the New York Central R. R. is scheduled to run from Chicago to New York, a distance of 1003 miles, in 20 hours. What is the average speed of this train in (a) miles per hour? (b) kilometers per hour? (c) feet per second?

2. The circumference of a cylindrical vessel is 20 cm. Find (a) the radius; (b) the cross sectional area.

3. The height of the cylindrical vessel (problem 2) is 20 cm. Find (a) its lateral area; (b) its volume.

4. What is the capacity of the vessel (problem 3) in (a) cc? (b) in quarts, liquid measure?

5. A cylindrical vessel when empty weighs 1760 g; full of water, it weighs 3330.8 g. The height of the cylinder is 20 cm. What is its radius?

6. The spherical bulb of a thermometer is 0.6 cm in diameter (inside measurement). It is filled with mercury the mass of which is 13.6 g per cm^3. How many grams of mercury are there in the bulb?

7. Twelve bicycle balls, each having a diameter of 1 cm, are dropped into a liquid the density (mass per cm^3) of which is 0.8 g per cm^3. Find (a) the volume of the liquid displaced; (b) the number of grams of liquid displaced.

CHAPTER II. **8.** A rectangular tank placed in a horizontal position is filled with water. It is 4 ft. high, 4 ft. wide, and 10 ft. long. What is the average pressure in lb./sq. ft. (a) on one end? (b) on one side?

9. What is the total force exerted by the water on the end and side of the tank (Ex. 8)?

10. An enclosed cylindrical metal tank, diameter 6 ft. and height 10 ft., is placed in an upright position. Into an opening in the top there is fitted a vertical tube 20 ft. long. Tank and tube are filled with water. What is the pressure on the bottom of the tank in lb./sq. ~~in.~~?

11. What is the force exerted by the water on the bottom of the tank (Ex. 10) in lb.~~/sq. ft.~~?

12. A swimming tank with a sloping bottom is 30 ft. wide and 80 ft. long. The water is 4 ft. deep at one end and 10 ft. at the other. What is the pressure in lb.; sq. ft. at the center of area of the bottom?

13. What is the force exerted by the water on the bottom of the tank (Ex. 12)?

14. What is the force on each end of the tank (Ex. 12)?

15. A pressure of 338 lb./sq. in. is required to force water to the top of the Woolworth Building, New York City. What is the height of the building?

16. A cubical vessel 12 cm on each edge is filled with a given liquid. The force exerted on one side of the vessel by the liquid is 11,750.4 g. What is the liquid?

17. A mass of lead having a volume of 1 cu. ft. and a specific gravity of 11.3 is lowered into water. What is the weight of the lead in water?

18. A hydraulic press is operated by water pressure from the city mains. The diameter of the piston of the press is 8 in. The load lifted is 3519 lb. What is the water pressure in lb./sq. in.?

19. A stone which weighs 9 lb. in air and 7.5 lb. in water is dropped into a vessel containing kerosene. How many pounds of kerosene does it displace?

20. A rod of uniform cross sectional area and weighted at one end sinks to a depth of 30 cm in water and to a depth of 40 cm in another liquid. What is the specific gravity of the liquid?

CHAPTER III. **21.** When on a certain day an aneroid barometer gives a reading equivalent to 76 cm at the bottom of the Woolworth Building, New York City, what will be the reading if it is carried to the top, the height being 780 ft.?

22. Over what height may water be siphoned on a mountain side half a mile above sea level, assuming that an ascent of 90 ft. is equivalent to a fall in the barometric column of 0.1 inch?

23. When the barometric pressure is 73 cm what is the maxi-

mum length of the short end of a siphon when used for sulphuric acid?

24. A given mass of gas has a volume of 1000 cm^3 when subjected to a pressure of 980,000 dynes per cm^2. What volume will the gas occupy if the pressure were decreased to 200 grams per cm^2, the temperature remaining constant?

25. What fractional part of the air originally in the receiver of an air pump still remains when the pressure within the receiver is equivalent to 1.2 cm of mercury, assuming that the air originally in the receiver was under a pressure of one atmosphere?

26. A gas bag filled with air is attached to a sinker and lowered into a lake. What will be the reduction in volume when the bag is 20 ft. below the surface?

27. The opening of a cubical metal tank 10 ft. on each edge is closed when the barometer stands at 30 in. Shortly thereafter the barometer drops to 29 in. What is the total outward force exerted on one side of the tank due to the difference of pressure within and without?

28. An air bubble occupies 10 cm^3 when submerged 8 meters below the surface of a pond. Assuming that the atmospheric pressure at the time is 1000 g/cm^2, what will be the buoyant force exerted on the bubble just as it reaches the surface?

29. A gas cylinder 5 ft. in length and 1 sq. ft. in cross sectional area contains gas under a pressure of 294 lb./sq. in. What volume would this gas occupy under normal atmospheric pressure?

30. Under standard conditions the density of air at the surface of the earth is 0.00129 g/cm^3 and the pressure which it exerts is 1033.3 g/cm^2. If the atmosphere were homogeneous (of uniform density) how many miles high would it be?

CHAPTER IV. **31.** A light gas of a given density and under a pressure of one atmosphere is enclosed in a porous vessel. The vessel is placed in a gas whose density is 25 times as great and whose pressure is also one atmosphere. What is the relative rate of diffusion of the two gases?

32. The height to which a liquid will rise in a capillary tube is directly proportional to the surface tension and inversely proportional to the radius of the tube and to the density of the liquid. Given two tubes *A* and *B*. Given also two liquids, *C* and *D*. The radii of the two tubes are equal. Assume that the densities of the two liquids are equal. The surface tension of liquid *C* is 66 g/cm and that of liquid *D* 44. Tube *A* is inserted into liquid *C* and tube

B into liquid D. What are the relative heights to which the liquids will rise in A and B?

33. Suppose that the surface tensions of C and D (Ex. 32) are the same, but that the specific gravity of C is 1.2 and that of D 0.75. What are the relative heights to which the liquids will rise in A and B?

34. Two spherical mercury globules, A and B of radii 1 mm and 2 mm respectively are brought into contact. They unite and form one globule, C. (a) What are the volumes of A and B? (b) What is the volume of C?

35. What is the radius of globule C (Ex. 34)?

36. (a) How does the area of globule C compare with the sum of the areas A and B? (b) Can you give any reason why the two globules are drawn together and unite in one when they come in contact?

37. A rod 5 m long and 1 cm^2 in cross section is supported in a vertical position by its upper end. A weight of 600 kgm is attached to the lower end producing a stretch in the rod of 3 mm. Find (a) the stress, (b) the strain.

38. A vertical rod 20 ft. in length and 1 in. in diameter is stretched by a certain weight 0.72 in. What stretch will be produced by the same weight in a rod of the same material and diameter, 30 ft. in length?

CHAPTER V. **39.** A stone thrown vertically upward just reaches the top of a 100 ft. smokestack. (a) With what speed did the stone leave the hand? and (b) what time did it take it to ascend?

40. A train coming into a station has at a given instant a velocity of 10 ft./sec. Five seconds later it has come to rest. (a) What was the negative acceleration? (b) How far did it move during the 5 seconds?

41. An electric car starts from rest and in 20 seconds it has acquired a speed of 30 mi./h. (a) What is the positive acceleration? (b) What distance did it travel during the 20 sec.?

42. A stone is thrown vertically downward from a cliff 800 ft. high with a velocity of 10 ft./sec. In what time will the stone reach the bottom of the cliff?

43. A body is dropped from the top of a tower 300 ft. in height. At the same instant another body is projected vertically upward from the base of the tower with a velocity of 300 ft. per second. At what point above the ground will the two pass each other?

44. A flag is hoisted to the masthead of a vessel a distance of 50 ft., during which time the vessel moves 100 ft. Find the magnitude and direction of the actual velocity of the flag.

45. A man riding on an electric car, which is running at the rate of 10 miles per hour, throws off a package at right angles to the track with a velocity of 10 ft. per second. Find the magnitude and direction of the velocity of the package.

46. Resolve a force of 100 lbs. into two components acting at an angle of 60° with each other, the angle which R makes with one of the components being 20°.

47. The wind is blowing northeast with a velocity of 20 miles per hour. Resolve this velocity into two components, one to the northward and one to the eastward.

48. A weight of 100 lbs. is supported by two ropes fastened to a beam, one rope making an angle of 30° with the beam, and the other 60°. Find the force exerted on each rope.

49. An ounce bullet is shot from a 9 lb. gun with a velocity of 1008 ft. per second. Find the velocity of the gun's recoil.

50. A boy throws a stone weighing 8 oz. with a velocity of 20 ft. per second. Compare the momentum of the stone with that of a freight car weighing 20 tons and moving with a velocity of 3 miles per hour.

51. Find the length of a pendulum that makes 90 vibrations per minute, giving the result in (a) meters; (b) feet.

52. A weight attached to a string is suspended from an upper story of a building. The weight swinging as a pendulum just clears the ground. It makes 10 vibrations per minute. Find the length of the string.

CHAPTER VI. **53.** What work will be required to hoist a load of ore from a shaft 510 ft. deep. under the following conditions? Weight of ore 2 tons; weight of car 600 lb.; weight of hoisting cable 10 lb. per yard.

54. An auto truck hauls a load weighing 3.5 tons along a level road a distance of 1.5 miles. The resistance offered by the road and the friction in the truck amounts to 5 lb. per hundredweight of the load. What amount of work was done?

55. A force of 100 lb. acts through a space of 100 ft. Find the work in (a) foot-poundals; (b) foot-pounds.

56. A force of 100 g acts through a space of 100 cm. Find (a) the work in ergs; (b) joules; (c) gram-centimeters.

57. The lower end of a ladder 30 ft. long stands on the ground at a distance of 6 ft. from the building against which the upper

end rests. How much work is done against gravity in carrying 100 lb. to the top of the ladder?

58. A team draws a load weighing 2 tons to the top of a hill the elevation of which is 300 ft. and the length of the incline 1 mi. The average force required to overcome frictional resistance is 300 lb. Find (a) the amount of work done against gravity, (b) against friction, (c) the total amount done by the team.

59. Forty-nine joules of work were done by a force acting through a distance of 2 m. Find the force in (a) dynes; (b) grams of force.

60. Find the horse power of an engine that can raise 10 tons to a height of 10 ft. in 10 seconds. What is the rating of the engine in kw?

61. A windmill pumps 125,000 gallons of water from a well 50 ft. deep in 10 minutes. Find its horse power.

62. It is estimated that 700 tons of water passes over Niagara Falls per minute. The distance which it falls is 160 ft. Compute the horse power of Niagara Falls.

63. The wind drives a boat at the rate of 10 miles an hour against an average resistance of 500 lb. Compute the h.p. furnished by the wind.

64. A mass of 10 grams has a velocity of 10 cm per sec. Find its kinetic energy in (a) ergs; (b) gram-centimeters.

65. A mass of 10 lb. has a velocity of 10 ft. per sec. Find its kinetic energy in (a) foot-poundals; (b) foot-pounds.

66. A uniform bar 10 ft. in length and weighing 2 lbs. to the foot is used as a lever of the first class. The fulcrum is placed 2 ft. from the end. Taking into account the weight of the lever, what downward force will be required at one end to overcome a resistance of 790 lbs. at the other?

67. A given bicycle wheel is 88 inches in circumference. The driving sprocket is 8 inches in diameter, the small sprocket 2 inches. How many times will the rider have to move his feet up and down in going a mile?

68. The axle of a windlass is 4 inches in circumference; the crank by which the windlass is turned is x inches. A force of 10 lbs. applied to the end of the crank supports 100 lbs. on the rope passing around the axle. Find the value of x.

69. The pilot wheel on a boat is 4 ft. in diameter; the axle 6 inches in diameter. What force must the pilot apply to the wheel in order to steer the boat, assuming that the steering resistance against the motion of the wheel is 360 lbs.?

70. A safe weighing 2 tons is to be raised into the end of a truck 3 feet above the ground by sliding it up a plank 16 feet long. If the average value of the frictional resistance between the safe and the plank is 80 pounds what is the smallest force required?

71. A barrel weighing 150 pounds is rolled up a 12-foot plank from the ground to a wagon 4 feet high. (a) Where on the barrel should the pushing force be applied to secure the greatest mechanical advantage? Why? (b) If the barrel were to slide without friction what force would be required? (c) How much work is done?

72. A 500 lb. weight rests upon a level sidewalk. What horizontal force is required to move the weight, the coefficient of friction being 0.4?

73. A 12 foot plank rests with one end on the ground and the other on a truck 3 ft. high. (a) What is the theoretical mechanical advantage of the incline formed by the plank? (b) Neglecting friction, what force parallel to the plank will be required to keep a 200 lb. box from slipping down the plank?

74. If the coefficient of friction between the box and the plank (Ex. 73) is 0.2, what force will be required to slide the box to the top of the incline?

75. A jackscrew is used to lift a load of 6 tons. The bar used to turn the screw is 3 ft. The pitch is 0.2 in. The efficiency of the screw is 30 per cent. (a) What is the theoretical mechanical advantage? (b) What is the actual mechanical advantage?

76. What force was required to lift the weight of 6 tons (Ex. 75)?

77. A force of 20 lb. is applied to the rope of a pulley system through a space of 6 ft. in order to raise a given weight through a distance of 6 in. The total resistance overcome is 180 lb. What is the efficiency of the system?

CHAPTER VII, HEAT. **78.** Find the value of the following temperatures on the Centigrade scale: (a) the temperature of the human body (98.6° F.); (b) the normal temperature of a living room (68° F.); the temperature of a cold day in winter (— 13° F.).

79. Find the value of the following temperatures on the Fahrenheit scale: (a) melting point of tungsten; (b) boiling point of glycerine.

80. What is the difference in Fahrenheit degrees between 10 degrees below freezing point Centigrade and 10 degrees below zero Fahrenheit?

81. At what temperature is the C reading twice the F reading?

82. At what temperature is the F reading twice the C reading?

83. A mass of 50 g of a substance at 100° C is dropped into 100 g of water at + 4° C, contained in a copper calorimeter having a mass of 60 g. The temperature of the water changes to + 10° C. What is the specific heat of the substance?

84. A mass of 340 g of lead is heated to 90° C and then dropped into 300 g of water contained in an aluminum calorimeter of mass 120 g. The final temperature of the water is 26° C. Find the initial temperature of the water.

85. A mass of 300 g of metal at 99° C is dropped into a hole in a block of ice at 0° C. The hole is immediately covered with another block of dry ice. A total mass of 33.5 g of ice is melted by the metal. Find the s.h. of the metal.

86. A metal wire 7.7 ft. in length at 0° C elongates 0.15 in. when heated to 90° C. (a) What it its coefficient of expansion? (b) What is the metal?

87. A certain distance measured with a steel tape was found to be 1025 ft. The measurement was made at 24° C. The tape was correct at 16° C. What was the correct distance measured, the coefficient of expansion of the tape being 0.000012?

88. What is the change in volume of a rectangular block of iron $3 \times 4 \times 10$ in. if the temperature changes from 44° F. to 116° F, the linear coefficient of expansion being 0.000012 per degree Centigrade?

89. A flask made of glass, linear coefficient of expansion 0.000008, is calibrated to hold 1000 cc at 0° C. What is its capacity at 100° C?

90. A solid displaces 500 cc when immersed in water at 0°, and 503 cc when immersed in water at 30° C. What is the cubical coefficient of expansion of the solid?

91. A cylinder is filled with gas under a pressure of one atmosphere. A piston within the cylinder stands at a height of 400 cm at 20° C. If the pressure be kept constant, what will be the height of the piston at 100° C?

92. If the piston (Ex. 91) be fixed so that the volume remains constant, what will be the pressure in atmospheres at 100° C?

93. How many B.t.u. will be required to change 10 lb. of ice to steam at 212° F?

94. The walls of an ice chest have an area of 15,000 cm^2, and are made of cork 3 cm thick. How many kg of ice may be ex-

pected to melt in a day (24 hours) when the average outside temperature is 86° F?

CHAPTER VIII, HEAT. **95.** The burning of a pound of coal produces about 14,000 B.t.u. If this energy could be used in lifting coal out of a mine 500 ft. deep, how many tons would be hoisted out of the mine for every ton burned?

96. In drilling a hole in a block of iron of mass 1 kg, power is supplied at the rate of 0.5 h.p. for 1 minute. If 75 per cent of the heat generated goes to heat the iron, what will be the change in its temperature?

97. How many feet of illuminating gas will have to be burned in order to furnish the same amount of heat (B.t.u.) as that given out by 1 ton of semi-bituminous coal?

98. If the efficiency of a gas engine is 20 per cent, how many horse power hours of work can be developed from the combustion of 1 lb. of gasoline, the heat of combustion being 20,000 B.t.u. per lb.?

99. A copper bar is 80 cm in length and 4 cm^2 in cross sectional area. One end of the bar is kept in steam at 100° C and the other end in ice. (a) How many calories per second will be transferred from one end of the bar to the other? (b) How many joules? (c) How many grams of ice will be melted?

100. A lead bullet of mass 10 grams strikes a target with a velocity of 500 m per second. Assuming that 20 per cent of the energy heats the bullet at the instant of impact, what will be its rise in temperature?

CHAPTER XI, CURRENT ELECTRICITY. **101.** How long will a current of 0.5 ampere have to flow to deposit 0.1 gram of silver on a metal spoon?

102. How much will a metal plate be increased in weight if it be copper-plated by a current of 0.5 ampere flowing for 5 hours?

103. A Daniell cell furnishes a current of 0.1 ampere for 1 hour. How much will the zinc electrode be diminished in weight due to the passage of the current?

104. An ammeter was calibrated by putting it in series with a silver coulometer. A constant current is passed through both instruments for 1 hour. The reading of the ammeter during this time was 0.76 ampere; the amount of silver deposited in the platinum bowl of the coulometer, is 3.016 grams. What was the error in the ammeter reading?

CHAPTER XII, ELECTRICITY. **105.** Compare the resistance of 4 m

of German silver wire, radius 0.2 mm, with that of 2,700 cm of nickel wire, radius 0.3 mm.

106. What length of copper wire of a given radius will be required to furnish a resistance equal to that of 3.2 m of manganin wire of the same radius?

107. Find the resistance of 10 miles of iron wire having a diameter of 150 mils.

108. A copper wire is 10 ft. in length; its diameter is 0.025 in. What is its resistance in ohms at 20° C?

109. Find the resistance of 6 m of pure platinum wire, B. & S. gauge No. 40, at 32° F.

110. Three conductors have resistances of 2, 4, and 6 ohms respectively. Find the resistance of the system when connected in (a) in series; (b) in parallel.

111. Enumerate all the resistances that can be obtained from three coils of resistances 2, 4, and 6 ohms respectively, by the various ways in which they may be connected, all three coils being always in use.

112. Given three conductors, *A*, *B*, and *C*, having resistances of 2, 3, and 6 ohms respectively, and a battery, having an e.m.f. of 7 volts and an internal resistance of 3 ohms. Find (a) the current flowing through the battery when *A*, *B*, and *C* are connected in series across the terminals; (b) the *IR* drop over each conductor.

113. The conductors *A*, *B*, and *C* (problem 112) are in parallel. Find (a) the current through the battery; (b) the *IR* drop over each conductor.

114. A storage battery of constant e.m.f., consisting of five cells, has a total internal resistance of 0.5 ohms. Connected to its terminals is a wire of 10 ohms of resistance. The current flowing through the wire is 1 amp. Find (a) the terminal potential (fall of potential over external circuit); (b) internal drop of potential; (c) e.m.f. of the battery.

115. A battery having an internal resistance of 5 ohms has two resistances *A* and *B* connected in series to its terminals. The resistance of *A* is 4 ohms and that of *B* 6 ohms. A current of 2 amp. flows through the circuit. Find (a) the fall of potential over the external circuit; (b) the e.m.f. of the battery.

116. A voltmeter having a resistance of 100 ohms is connected in parallel across *A*, problem 115. Find (a) the resistance of the entire circuit; (b) the current through *B*.

117. What is the reading of the voltmeter (Prob. 116)?

118. Five dry cells, each having an e.m.f. of 1.5 volts, and an

internal resistance of 0.1, 0.2, 0.3, 0.4, 24 ohms respectively, are connected in series. (a) What current will the five cells furnish through an external resistance of 5 ohms? (b) What current will be furnished through the same resistance if the last cell be cut out?

119. A current of 1.2 amp. under a pressure of 110 volts flows through an incandescent lamp for 5 hr. (a) Find the power expended in watts. (b) Find the energy expended in watt-hours.

120. What power, in watts, is used by a conductor, (a) which takes 2 amp. at 110 volts? (b) which takes 2 amp. through a resistance of 55 ohms?

121. How much energy, in ergs, is used by a 55-watt lamp burning 10 min.?

122. A bank of incandescent lamps takes 10 amp. at 110 volts, 80 per cent of the energy received being given off as heat. Find the heat given off in 1 hr. in (a) joules; (b) calories.

123. Two wires, A and B, of the same material and length are connected in series. The diameter of A is to that of B as 2:1. The resistance of A is 2 ohms. The applied e.m.f. is 10 volts. Find the heat in calories generated in B during 10 sec.

124. An electrolytic cell for the purification of copper has a resistance of 0.01 ohm. How much heat will be developed in this cell per minute at a time when 0.987 gram of copper is being deposited per minute?

125. A storage battery has an e.m.f of 60 volts and an internal resistance of 0.8 ohm. When it is being charged a current of 20 amp. is used. (a) What is the applied e.m.f., and (b) what is the rate at which energy is stored in the battery?

CHAPTER XIII. **126.** Consider a simple two-pole generator, such as shown in Fig. 400, page 515. The armature, which consists of a single loop of wire, makes 360 r.p.m. in a magnetic field which consists of 200,000 lines of force. Assuming that during each revolution each of the wires A and C cuts the entire number of lines twice, what e.m.f. is developed?

127. A dynamo armature has 50 coils of 5 loops each and makes 600 r.p.m. Each wire cuts 2,000,000 lines of force twice per revolution. What is the e.m.f. generated?

128. A d.c. generator operating at 500 volts furnishes to a factory a current of 50 amperes through a line having a resistance of 2 ohms (a) How much power is developed by the generator? (b) What is the IR drop in the line? (c) At what voltage does the factory receive its current?

129. A generator, the voltage across the terminals of which is 220 volts, is connected to a distant motor by wires having a resistance of 0.5 ohm. The loss of voltage (*IR* drop) in the wires is 20 volts. (a) How much current is supplied to the motor? (b) What is the difference of potential of the motor terminals?

130. The voltage across the terminals of a d.c. motor is 205 volts, and the current is 20 amperes. If the internal resistance is 0.5 ohm, what is the counter e.m.f. of the motor?

131. A generator having an internal resistance of 0.5 ohm is connected by wires with a resistance of 1.5 ohms, to a place where the power is used. At this place a voltmeter shows 114 volts, and an ammeter 25 amperes. What is the e.m.f. of the generator?

132. If it requires 100 h.p. to drive a generator producing 137 amperes at 500 volts terminal potential difference, find the efficiency of the machine.

133. Find the horse power required to operate 150 incandescent lamps each taking 0.45 amperes at 110 volts.

134. Find the current used by an 8 h.p. 500 volt motor if its efficiency is 90 per cent.

CHAPTER XIV. **135.** A transmission line has a resistance of 2 ohms. (a) How much power is lost in heat when the current is 10 amperes? (b) What is the heat loss when the current is 100 amperes?

136. Find the current used if 110 kw. are to be transmitted (a) at 110 volts; (b) at 2,200 volts.

137. For the same weight of copper conductors and the same percentage drop in the lines, compare the distance to which the same power could be transmitted, using the two voltages 110 and 2,200 as given in problem 136.

138. For the conditions given in problem 137, compare the heat losses in the lines.

139. The electric power of 50 kw. is delivered to a transmission line of resistance 5 ohms. (a) What power is wasted in the line if the current is delivered under a pressure of 10,000 volts? (b) 1000 volts?

140. A transformer receives electric energy under a pressure of 550 volts and raises it to 66,000 volts. The efficiency of the transformer is 97 per cent. What is the high voltage current when the low voltage current is 600 amperes?

CHAPTERS XVI AND XVII. **141.** A person approaching a large building at night stamps his foot on the pavement, and 0.8 sec.

later hears the echo. How far is the person from the building, assuming the temperature to be 68° F.

142. A cliff 460 ft. distant returns an echo to an observer in 0.8 sec. Find the temperature.

143. Two large buildings are 346.4 ft. apart. The air is still and the temperature is 24° C. A pistol is fired at a distance of 118.8 ft. from one of the buildings and 227.6 ft. from the other. How long before the man who fired the pistol will hear the third echo?

144. The density of the air is 14.4 times that of hydrogen. Find the velocity of sound in hydrogen at 0° C, the pressure and coefficient of elasticity in both cases being the same.

145. The velocity of sound in aluminum is 5200 m per sec. Find its coefficient of elasticity.

146. Find the time of reverberation for an empty room $12 \times 20 \times 30$ ft., assuming that the average absorption coefficient of the ceiling, walls, and floor is 0.36 per sq. ft.

147. Find the total absorbing power of a room if the time of reverberation is 0.5 sec. and the volume is 50,000 cu. ft.

148. The whistle of a locomotive has a frequency of 300 vibrations per second. When the temperature is 20° C, what will be the apparent frequency if it is moving (a) away from the observer with a speed of 60 mi. per hr.? (b) toward the observer at the same rate?

149. A sounding body makes 100 vibrations per sec. Find the wave length of the disturbance in air at (a) $+20°$ C; (b) $-20°$ C.

150. A tuning fork gives off waves in air 150 cm in length, at 16.7° C. Find (a) the frequency of the fork; (b) its period.

151. Middle *C* is assumed to have 256 vibrations per sec. Compute the vibration frequency of each of the other three notes of the major chord.

152. A cylindrical glass tube is placed vertically in water, and its length is adjusted until it responds to a fork making 256 vibrations per sec., when the temperature is 20° C. Find the length of the resonance tube in feet.

153. A closed resonance tube, 11.4 in. in length, responds to a fork making 300 vibrations per sec. Find the temperature.

154. A given string stretched by a force of 1 lb. makes 200 vibrations per sec. (a) How will the frequency be affected if the stretching force is increased to 4 lb.? (b) How will the pitch be affected?

155. If the stretching force upon a certain string is 500 grams, and it sounds middle *C*, what will be the force necessary to tune it to *D* of the natural scale?

156. A wire 3 ft. long and stretched with a force of 9 lb. makes 300 vibrations per sec. Another wire of the same material and size is 4 ft. long, and is stretched with a force of 16 lb. Find the vibration frequency of the second wire.

157. Two wires of the same material and the same size have the same pitch. One is stretched with a force of 4,900 grams and is 80 cm. long; the other is stretched with a force of 8,100 grams; how long is the second wire?

158. Determine the length of an open organ pipe that is in unison with the note of the first black key above *A* of the piano when the temperature of the air is 25° C?

159. What is the vibration frequency of an open organ pipe 32 ft. long when the temperature is 24° C?

160. The frequency of the lowest continuous sound which the ear can perceive as a definite note is about 20 per sec.; the limits for the highest, is about 40,000 per sec. Find the corresponding wave lengths in ft. at 0° C.

CHAPTERS XVIII AND XIX. **161.** A beam of light falls upon a plane mirror which lies in a horizontal position on the ground 20 ft. distant from the vertical wall of a building. The reflected ray strikes the wall at a point 10 ft. from the ground. Find the angle of incidence of the light on the mirror.

162. A 25-watt lamp is rated at 1.1 watts per c.p. What intensity of illumination in foot-candles is produced at a distance of 3 ft. from the lamp?

163. A 32 c.p. electric lamp is 3 ft. from a book. How far from the book must a 300 c.p. lamp be placed in order to produce the same intensity of illumination?

164. An incandescent lamp is rated at 0.7 watt per c.p. What size of lamp, rated in watts, is needed to produce an illumination of 8 foot-candles at a distance of 3 ft.

165. By means of a photometer it is found that two lamps produced the same intensity of illumination at a point 120 cm from the first lamp and 80 cm from the second. If the candle power of the first lamp is 24, what is the candle power of the second?

166. Taking Michelson's value for the velocity of light in air as 2999×10^7 cm/sec., find the velocity of light in glass, the index of refraction of which is 3/2.

167. An object 9 feet from a spherical concave mirror gives a

real image 12 feet from the mirror. Find radius of curvature of the mirror.

168. A concave mirror has a focal length of 5 cm. What is the size of the image when an object 5 cm high is 30 cm from the mirror?

169. An object 10 cm in height is placed 15 cm in front of a convex mirror the radius of which is 6 cm. Find the size of the image.

170. (a) What kind of a mirror will produce an erect image of an object one-half of its natural size when the object is 10 in. and the image is 5 in. from the mirror? (b) What is the radius of curvature of the mirror?

171. The focal length of a bi-convex lens is 1 ft. Where must an object which is 6 in. in length be placed so that the image will be 18 in. from the lens and real?

172. What is the size of the image (problem 171)?

173. The focal length of a bi-convex lens is 1 ft. Where must an object, which is 6 in. in length, be placed so that the image will be 18 in. from the lens and virtual?

174. What is the size of the image (problem 173)?

175. If an object 10 in. in length be placed 3 ft. from a bi-concave lens, the focal length of which is 1 ft., what will be the size of the image?

176. The length of the tube of a compound microscope is 12 cm; the focal length of the eyepiece is 0.4 cm. What must be the focal length of the objective to secure a magnification of 1500?

177. The objective of the Lick Observatory telescope has a focal length of 1500 cm. Find the focal length of an eyepiece that will give a magnifying power of 400.

178. If the nearest distance for distinct vision for a farsighted person is 35 in., what should be the focal length of the spectacles he would require for reading at a distance of 10 in.?

179. If the greatest distance of distinct vision for a myopic (nearsighted) eye is 5 in. what should be the focal length of the proper reading spectacles?

180. A picture 2×3 in. on a lantern slide is focused on a screen which is 30 ft. from the lantern. The image on the screen is 6×9 ft. What is the focal length of the projection lens?



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Kinetic theory of gases:
All gases are made up of minute
particles called molecules.
The molecules of a given gas are
all exactly alike in weight.
The molecules of a gas are in rapid
and constant motion.
Heat increases the speed of the
molecules and cold decreases the speed.

Gases
fluency
compressibility
High diffusion
All gases miscible
No definite surface
Not viscous
And to fill container

Liquids
1. fluency
2. low compressibility
3. low diffusion
4. All liquids not miscible
5. definite surface
6. high viscosity
7. Tend to take shape of cont

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