PMS/CSS

Everyday Science

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It's just an effort to merge all relevant data of Everyday Science in a single document, which will be used in the preparation of Competitive Examinations like PMS/CSS and other such exams. The primarily source of these information is mainly from internet.

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Notes

Contents

1. NATURE OF SCIENCE

Definition:
Science can be defined as study
"mainly concerned with the phenomenon of physical universe any or all of natural sciences or biological sciences."
Or
Science as the "the field of study which attempts to describe and understand the nature of the universe in whole or part."

Science is the faculty to reason out the how and why of the things as they occur in the phenomenal world or the objective world...Basically science is the study of laws of nature and man has developed science by observing.Infact this subject has completely transformed our power over nature and the world outlook. Development of the modern technology is directly the outcome of the development of modern science. Without the scientific revolution the industrial revolution would not have been possible.

It has raised the human activity level by significant observations in the various fields of human existence. Whether it's the exploration of human health, industrial progress, agrarian developments and modern communication technologies, the benefits gained from this major subject are enormous.Infact it would not be wrong to say that we are living in the age of science and is a dominant factor in our day to day existence.

2. CONTRIBUTIONS OF MUSLIM SCIENTISTS

MUHAMMAD BIN MUSA AL KHWARZIMI:

Made lasting contributions in the fields of Mathematics, Astronomy, Music, Geography and History. He composed the oldest works on Arithmetic and on Algebra. The oldest Mathematic book composed by him is "Kitab ul jama wat tafriq"
He is the first person who used zero and wrote"Hisab ul jabr Wal Muqabla" which is conceived to be an outstanding work on the subject which included analytical solutions of linear and quadratic equations.
In the field of Astronomy he compiled his own tables which formed the basis of later astronomical pursuits in both East and West. He also contributed in the field of geographical science by writing a noteworthy book KItab ul Surat al ard. In Arabic.
His book —kitab al Tarik" is also a memorable work regarding history.

AL BERUNI:

Born in Afghanistan Beruni made original important contributions to science. He is conceived to be the most prominent scientists of the Islamic world who wrote around 150 books on various significant subjects concerning human existence. These subjects include Mathematics, History, Archeology, Biology, Geology, Che m is try, Religion etc.
He discussed the behavior of earth, moon, and planets in his book "Qanoon Almasudi" which is also considered as an outstanding astronomical encyclopedia. He also discovered seven different ways of finding the directions of north and south and discovered mathematical techniques to determine exactly the beginning of the seasons.
Another notable discovery he made was that the speed of light is faster than sound .His wide range of scientific knowledge is also revealed through his books" kitab al saidana" and "kitab al jawahar" dealing with medicine and the types of gems their gravity respectively.
He was a prolific writer whose works showed his versatility as a scientist.

AL RAZI:

The famous philosopher and a notable surgeon of the Muslim world, Zakriya Al Razi was born in Ray near modern Theran Iran. His eagerness for knowledge lead him to the study of Alchemy and Chemistry, philosophy, logic, Mathematics and Physics. He was a pioneer in many areas of medicine and treatment of health sciences in general, and in particular he worked alot in the fields of paeditrics, obsterics and ophthalmology.
Al razi was the first person to introduce the use of Alcohol for medical purposes and opium for the objective of giving anesthesia to his patients.
In the field of ophthalmology too Al razi gave an account of the operation for the extraction of the cataract and also the first scientist to discover the effect of the intensity of light on the eye. The modern studies confirm his understanding on the subject thus making him a great physician of all the times.

ABU ALI IBN E SINA:

Endowed with great powers of absorbing and retaining knowledge this Muslim scholar also made valuable contributions to the field of science. He is considered to be the founders of Medicine and also added his great efforts to the fields of Mathematics, Astronomy, Medicinial Chemistry, Philosophy, Palae ontology and Music.
His most famous book is "Al Qannun" which brings out the features of human physiology and medicine.
Sina is also considered as a father of the science of Geology on account of his invaluable book on mountains in which he discussed matters relating to earth's crust and gave scientific reasons for earthquakes. He is the author of 238 books which are fine instances of his thoughts regarding various subjects in diverse ways.

JABIR BIN HAYAN:

Introduced experimental research in chemical science which immensely added its rapid development and made him the Father of Chemistry. He devised methods for preparation of important chemicals like hydrochloric acid, nitric acid, and white lead.

Jabir's work also deal with the refinement of metals ,preparation of steel, dyeing of cloth and leather, use of magnese dioxide in glass making, distillation of vinegar to concentrate acetic acid.
Jabir also explained scientifically two principle functions of chemistry, i.e., calcination, and reduction and registered a marked improvement in the methods of evaporation, sublimation, distillation and crystallization
He wrote more than 100 books which are one of the most outstanding contributions in the field of science especially the chemical science.

ABDUL HASSAN IBN AL HAITHAM:

One of the most outstanding Mathematicians, Physiologists, and Opticians of Islam. He contributed to the realms of medicine and philosophy. He wrote more than 200 scientific works on diverse subjects.
Haitham examined the refraction of light rays through transparent objects including air and water.
Infact he was the first scientist to elaborate two laws of reflection of light
He made a number of monumental discoveries in the field of optics ,including one which locates retina as the seat of vision. His book on optics "Kitab Al Manazir" vividly shows his grip on the subject.
He constructed a pinhole camera and studied formation of images .Due to his noteworthy contributions he is regarded as one of the prolific Muslim scientists of all times.

OMAR AL KHAYAM:

He was an outstanding Mathematician and Astronomer. He was also known as a poet, philosopher and a physician. He travelled to the great centers of learning of the era i.e. Samrakund, Bukhara, and Ispahan.He classified many algebraic equations based on their complexity and recognized thirteen different forms of cubic equation. He also classified algebraic theories of parallel lines. On the invitation of Sultan Jalal-ud- Din, he introduced the Jilali calendar which has an error of one day in 3770 years. He also developed accurate methods for determination of gravity
as a poet too, he is known for his Rubaiyat.He made great contributions in the development of mathematics and analytical geometry which benefitted Europe several years later.

NASIR UD DIN TUSI:

Al tusi was one of the greatest scientists, Mathematicians, Astronomers, Philosophers, Theologians and physicians of his time. He was a prolific writer and wrote many treatises on varied subjects like Algebra, Arithmetic, Trignometry, Geometery, Logic, Met aphy sics, medicine, ethics and Theology.
He served as a minister of Halaku Khan and persuaded him to establish an observatory and library after the destruction of Baghdad. He worked at the observatory and prepared precise tables regarding the motion of the planets. These are also known as "Tables of Khan"

ZIA UD DIN IBN BAITAR:

Was a famous botanist and pharmacopeias of middle ages. Because of his intensive travels, he was able to discover many plant species. He wrote many books regarding his field of specialty and is always considered as a prominent scientist among his Muslim counterparts

3. IMPACT OF SCIENCE ON SOCIETY

Science is the organization of knowledge in such a way that it commands the hidden potential in nature. This hidden potential is surfaced out by the subject of science through the process of understanding. Science has proved to be of enormous beneficial nature .It has made lasting impact on regarding each and every field of human existence. Whether it is concerned with our day to day lives or whether it is related with the various modern developments which have resulted in elevating the living standards of the individuals.
The significant contributions which the study of this subject has made are enumerated below.

SCIENCE AND HUMAN ATTITUDE:

The various noteworthy scientific advances have helped the individuals in raising up of their self confidence .This subject has enabled the human beings to control and modify their needs and requriements.With greater understanding of the scientific phenomena human beings have now become more confident about the environmental issues as compared to the people in the past.Infact science has promoted and paved the way for the independent and logical thinking.

SCIENCE AND HUMAN HEALTH:

Before the development of modern medicinal factors, a large number of people used to lose their precious lives because of the unavailability of the sources and medicines for a proper health care. With the advancements of science now the human life expectancy rate has increased as the various modern developments in the field of health care has helped in warding off the dangerous diseases...
The revolutions in surgery and medicine the infectious diseases like small pox, malaria, typhoid etc. have been eradicated. Thus science has improved the health standards of the people.

SCIENCE AND TRAVEL:

People used to travel on foot before the inventions of automobiles,aeroplanes and steam engines. They also used animal carts and camels for the purpose of moving from one place to another. However, the modern scientific inventions have proved to be of great significance as it has added speed to the area of travel. The quick means of transportation have decreased the distances and are a source of saving time. In fact it would not be wrong to regard that these inventions have added much peace to the lives of the modern men.

SCIENCE AND COMMUNICATION:

Science has also played a significant part in the development of the modern communication technology. Earlier people were living in isolation because of the slow means of communication. Now the well developed, efficient media have made it possible to communicate with each other more rapidly and quickly. The impact of mass media is enormous. The use of computers and televisions has made the world a global village where an event in one part of the world leaves an influence on the other.

DEMERITS OF SCIENCE:

Every invention of science has got its own merits and demerits. The most serious invention that science has contributed to is the development of the weapons of mass destruction like the atom and nuclear bombs. The recent wars have greatly showed that how much destruction can be brought about with the use of these lethal weapons. In fact these modern inventions of science have resulted in the elevation of the anxiety and unrest in the modern societies.

Another notable demerit which the study of this subject has lead to the rise in the environmental deterioration. Day by day the pollution factor is increasing which has proved to be very toxic and harmful for the human health. Not only the human health it is also proving fatal for the animals as well as the existing plants.

The rapid developments of science and industrialization have also divided the world. The developed and the undeveloped. This division has lead to a widening gap between the status and the living standards of people. There is economic disparity which has also given rise to class distinction

4. UNIVERSE

The BIG BANG THEORY about the universe is the most widely acceptable theories with regard to the origin of the universe. According to the big bang, the universe was created sometime between 10 billion and 20 billion years ago from a cosmic explosion that hurled matter and in all directions. All the galaxies were formed from this matter.
Observations of these galaxies show that they are still moving apart from each other .The universe is expanding Some scientists have suggested another theory as "steady theory:" to explain the process of the evolution of the universe. However the general notion on which all scientists agree is the theory of Big Bang.

Steady theory is the theory about the universe and the observations by the astronomers have shown that the galaxies are moving away from each other and the universe seems to be expanding. The theory shows that the new matter is always being created to fill the space left by this expansion. The new matter moves apart and forms galaxies which continue to move apart. This means that the universe always look exactly the same. It has no beginning or end but in a steady state. However many observations have suggested that the universe has not always looked like the same.

THE FUTURE OF UNIVERSE:

At present the universe is expanding but the astronomers have questioned that whether or not this expansion will continue .Certain observations which have been made in this regard is that one possible ending of the universe will be the "big crunch‖.
The galaxies and other matter may be moving apart but their motion is restrained by their mutual gravitational attraction. If there is a sufficient matter in the universe gravity will eventually win and begin pulling the galaxies together again causing the universe to experience a reverse of the big bang i.e., the BIG CRUNCH.
However there is a possibility that there is not enough matter in the universe for the big crunch to happen. This means that if it happens then the universe will continue to expand forever.

5. GALAXY

Galaxy is a huge number of stars grouped together. The term galaxy can also be described as a collection of dust, gas and stars measuring thousands of parsecs across.Galzxy contains 10000million stars and looks like a disc with a fat centre and spiral arms. From the front it looks like a convex lens's

Classes of galaxy:

Two broad classes of galaxy are there.
1. Elliptical
2.Spiral
the spiral galaxies are further sub divided into normal which constitutes of majority of spirals and barred spirals. Barred spirals have their centre in the form of the bar.
The elliptical galaxies range from E 0 to E 7 from an almost spherical shape to a flattened disc.

Milky Way:

Our galaxy is a spiral galaxy about 30,000 parsecs across. There are more than 200 billion stars in the galaxy. Its disc appears as a faint white band that is responsible for dividing the white sky at the night into two. The name of our galaxy is Milky Way.

The galaxy has three spiral arms called the Orion, Perseus, and Sagittarius arms and the whole system is rotating in space. The sun revolves around the nucleus of the galaxy once in 225 million years. This duration is also called the cosmic year.

I -INTRODUCTION:

Milky Way, the large, disk-shaped aggregation of stars, or galaxy, that includes the Sun and its solar system. In addition to the Sun, the Milky Way contains about 400 billion other stars. There are hundreds of billions of other galaxies in the universe, some of which are much larger and contain many more stars than the Milky Way.

The Milky Way is visible at night, appearing as a faintly luminous band that stretches across the sky. The name Milky Way is derived from Greek mythology, in which the band of light was said to be milk from the breast of the goddess Hera. Its hazy appearance results from the combined light of stars too far away to be distinguished individually by the unaided eye. All of the individual stars that are distinct in the sky lie within the Milky Way Galaxy.

From the middle northern latitudes, the Milky Way is best seen on clear, moonless, summer nights, when it appears as a luminous, irregular band circling the sky from the northeastern to the southeastern horizon. It extends through the constellations Peruses, Cassiopeia, and Cepheus. In the region of the Northern Cross it divides into two streams: the western stream, which is bright as it passes through the Northern Cross, fades near Ophiuchus, or the Serpent Bearer, because of dense dust clouds, and appears again in Scorpio; and the eastern stream, which grows brighter as it passes southward through Sputum and Sagittarius. The brightest part of the Milky Way extends from Sputum to Scorpio, through Sagittarius. The center of the galaxy lies in the direction of Sagittarius and is about 25,000 light-years from the Sun (a light-year is the distance light travels in a year, about 9.46 trillion km or 5.88 trillion mi).

II -STRUCTURE:

Galaxies have three common shapes: elliptical, spiral, and irregular. Elliptical galaxies have an ovoid or globular shape and generally contain older stars. Spiral galaxies are disk-shaped with arms that curve around their edges, making these galaxies look like whirlpools. Spiral galaxies contain both old and young stars as well as numerous clouds of dust and gas from which new stars are born. Irregular galaxies have no regular structure. Astronomers believe that their structures were distorted by collisions with other galaxies.

Astronomers classify the Milky Way as a large spiral or possibly a barred spiral galaxy, with several spiral arms coiling around a central bulge about 10,000 light-years thick. Stars in the central bulge are close together, while those in the arms are farther apart. The arms also contain clouds of interstellar dust and gas. The disk is about 100,000 light-years in diameter and is surrounded by a larger cloud of hydrogen gas. Surrounding this cloud in turn is a spherical halo that contains many separate globular clusters of stars mainly lying above or below the disk. This halo may be more than twice as wide as the disk itself. In addition, studies of galactic movements suggest that the Milky Way system contains far more matter than is accounted for by the visible disk and attendant clusters—up to 2,000 billion times more mass than the Sun contains. Astronomers have therefore speculated that the known Milky Way system is in turn surrounded by a much larger ring or halo of undetected matter known as dark matter.

III -TYPES OF STARS:

The Milky Way contains both the so-called type I stars, brilliant, blue stars; and type II stars, giant red stars. Blue stars tend to be younger because they burn furiously and use up all of their fuel within a few tens of millions of years. Red stars are usually older, and use their fuel at a slower rate that they can sustain for tens of billions of years. The central Milky Way and the halo are largely composed of the type II population. Most of this region is obscured behind dust clouds, which prevent visual observation.

Astronomers have been able to detect light from this region at other wavelengths in the electromagnetic spectrum, however, using radio and infrared telescopes and satellites that detect X rays (see Radio Astronomy; Infrared Astronomy; X-Ray Astronomy). Such studies indicate compact objects near the galactic center, probably a massive black hole. A black hole is an object so dense that nothing, not even light, can escape its intense gravity. The center of the galaxy is home to clouds of antimatter particles, which reveal themselves by emitting gamma rays when they meet particles of matter and annihilate. Astronomers believe the antimatter particles provide more evidence for a massive black hole at the Milky Way's center.

Observations of stars racing around the center also suggest the presence of a black hole. The stars orbit at speeds up to 1.8 million km/h (1.1 million mph)—17 times the speed at which Earth circles the Sun—even though they are hundreds of times farther from the center than Earth is from the Sun. The greater an object's mass, the faster an object orbiting it at a given distance will move. Whatever lies at the center of the galaxy must have a tremendous amount of mass packed into a relatively small area in order to cause these stars to orbit so quickly at such a distance. The most likely candidate is a black hole.

Surrounding the central region is a fairly flat disk comprising stars of both type II and type I; the brightest members of the latter category are luminous, blue supergiant. Imbedded in the disk, and emerging from opposite sides of the central region, are the spiral arms, which contain a majority of the type I population together with much interstellar dust and gas. One arm passes in the vicinity of the Sun and includes the great nebula in Orion. See Nebula.

IV -ROTATION:

The Milky Way rotates around an axis joining the galactic poles. Viewed from the north galactic pole, the rotation of the Milky Way is clockwise, and the spiral arms trail in the same direction. The period of rotation decreases with the distance from the center of the galactic system. In the neighborhood of the solar system the period of rotation is more than 200 million years. The speed of the solar system due to the galactic rotation is about 220 km/sec (about 140 mi/sec).

6. SOLAR SYSTEM

The solar system includes nine planets and sun being at the centre. All the planets revolve around the sun .The solar system also includes the asteroids, meteors and numerous comets. All of these travel around the sun in a particular orbit .The planets which are the significant part of the solar system namely,Mercury,venus,earth,mars,Jupiter,Saturn<ur anus,neptune and the Pluto.

All the theories about the formation of the solar system agree on two facts. One is that the age of solar system is 4.5billion years and secondly the planets were formed from the gas and dust within the vicinity of the sun. The gas and dust condensed into tiny bodies which eventually built up the present day planetary system.

MOTION OF THE PLANETS:

The planets perform two types of motions.
1. rotation at their axis is
2. revolution or the orbital motion along their orbits around the sun.

THE AGE OF THE SOLAR SYSTEM:

Meteoric evidence suggests that the solar system is 4530+20 million years old. And that was formed in less than 25 million years. The solar system is believed to be formed from a globe of gas and dust that consisted mainly of hydrogen

7. THE SUN

The sun is the most significant star for the existence and life of living beings on earth. The whole life on earth depends on this brightest object in the sky. The living things get their energy from sun and the appropriate distance of earth from sun maintains an appropriate temperature which is ideal for the survival of all the life present on earth.
The sun is the principle source of energy for all the planets of the solar systems .Through a constant stream of particles that flow outward from the sun, heat, radiation, light and UV rays are emitted.the UV rays that reach the earth from sun are considered to be the most harmful ones.Most of the UV and other high energy radiation are absorbed by the Ozone layer and the atmosphere of the earth.This stream of particles is called solar wind.

STRUCTURE OF THE SUN:

The visible surface of the sun is called photosphere which is a turbulent region and is 570 kilometer deep.
The layer outside the photosphere is chromospheres which is a broad layer and is several miles thick.
The outside layer of the sun is corona. Energy released by the sun passes from chromospheres to the corona and then to the outer space.

8. EARTH

The third farthest planet from the sun is earth. The earth rotates on its axis in about 24 hours. The diameter of earth is 12756 km.The earth completes its rotation along its axis in 23.9 hour and one tip along its orbit in 365 days,6 hours and 9 minutes and 9.4 seconds. The orbit of the earth is not a circle but an ellipse.

STRUCTURE OF THE EARTH:

The earth has four major structural components.
1. The Crust
2.The Mantle
3.The Outer core
4.The Inner Core.

THE CRUST: The upper most layer of the earth is called the crust. A solid and a complex layer of the lithosphere in both physical and chemical nature. The crust is composed of wide variety of rocks which are known as sedimentary rocks .The crust is further divided into ocean crust and continental crust.

THE MANTLE: According to the various kinds of scientific evidences the upper mantle of the earth is made up of silicate minerals. The temperature increases rapidly with depth in outer of the earth. Due to high temperatures the rocks start melting .These molten rocks form the basis of lava which erupt from oceanic volcanoes.

THE CORE: The core starts where marked physical and chemical changes occur across the vore mantle boundries.The outer core is thought to be made of mixture of melted iron and nickle.In the inner core the mixture is thought to be nickel and iron alloy.

ATMOSPHERE OF THE EARTH:

The chief gases in the atmosphere of the earth are,
Nitrogen 78.09%
Oxygen 20.95%
Argon 0.93%

The remaining 0.03% is made of carbon dioxide, small quantities of neon,helium,ozone and hydrogen and minute traces of krypton,methane,xenon and other gases.

Another important constituent of the atmosphere is water vapor which makes up 4 percent by volume and 3 percent by weight. Life is not possible without the atmosphere of the earth Oxygen present in the atmosphere is necessary for both animals and plants for respiration. Carbon dioxide is needed by the plants which use this gas in the process of photosynthesis. Water vapors present in the earth are necessary for the process of rain and other precipitation.
The atmosphere is a valuable natural source of many gases that are widely needed in industry, argon for the purpose of welding and oxygen is required for hospitals and other metal industries.
The earth's atmosphere aloe possesses the protective role against the dangerous UV rays and other high radioactive energy from space. The atmosphere of the earth absorbs these radiations which are a cause of various health hazards.

9. ATMOSPHERE

Atmosphere, mixture of gases surrounding any celestial object that has a gravitational field strong enough to prevent the gases from escaping; especially the gaseous envelope of Earth. The principal constituents of the atmosphere of Earth are nitrogen (78 percent) and oxygen (21 percent). The atmospheric gases in the remaining 1 percent are argon (0.9 percent), carbon dioxide (0.03 percent), varying amounts of water vapor, and trace amounts of hydrogen, ozone, methane, carbon monoxide, helium, neon, krypton, and xenon.

The mixture of gases in the air today has had 4.5 billion years in which to evolve. The earliest atmosphere must have consisted of volcanic emanations alone. Gases that erupt from volcanoes today, however, are mostly a mixture of water vapor, carbon dioxide, sulfur dioxide, and nitrogen, with almost no oxygen. If this is the same mixture that existed in the early atmosphere, then various processes would have had to operate to produce the mixture we have today. One of these processes was condensation. As it cooled, much of the volcanic water vapor condensed to fill the earliest oceans.

Chemical reactions would also have occurred. Some carbon dioxide would have reacted with the rocks of Earth`s crust to form carbonate minerals, and some would have become dissolved in the new oceans. Later, as primitive life capable of photosynthesis evolved in the oceans, new marine organisms began producing oxygen. Almost all the free oxygen in the air today is believed to have formed by photosynthetic combination of carbon dioxide with water. About 570 million years ago, the oxygen content of the atmosphere and oceans became high enough to permit marine life capable of respiration. Later, some 400 million years ago, the atmosphere contained enough oxygen for the evolution of air breathing land animals.

The water-vapor content of the air varies considerably, depending on the temperature and relative humidity. With 100 percent relative humidity, the water vapor content of air varies from 190 parts per million (ppm) at -40°C (-40°F) to 42,000 ppm at 30°C (86°F). Minute quantities of other gases, such as ammonia, hydrogen sulfide, and oxides of sulfur and nitrogen, are temporary constituents of the atmosphere in the vicinity of volcanoes and are washed out of the air by rain or snow. Oxides and other pollutants added to the atmosphere by industrial plants and motor vehicles have become a major concern, however, because of their damaging effects in the form of acid rain. In addition, the strong possibility exists that the steady increase in atmospheric carbon dioxide, mainly as the result of the burning of fossil fuels since the mid-1800s, may affect Earth`s climate.

Similar concerns are posed by the sharp increase in atmospheric methane. Methane levels have risen 11 percent since 1978. About 80 percent of the gas is produced by decomposition in rice paddies, swamps, and the intestines of grazing animals, and by tropical termites. Human activities that tend to accelerate these processes include raising more livestock and growing more rice. Besides adding to the greenhouse effect, methane reduces the volume of atmospheric hydroxyl ions, thereby curtailing the atmosphere`s ability to cleanse itself of pollutants.

The study of air samples shows that up to at least 88 km (55 mi) above sea level the composition of the atmosphere is substantially the same as at ground level; the continuous stirring produced by atmospheric currents counteracts the tendency of the heavier gases to settle below the lighter ones. In the lower atmosphere, ozone, a form of oxygen with three atoms in each molecule, is normally present in extremely low concentrations. The layer of atmosphere from 19 to 48 km (12 to 30 mi) up contains more ozone, produced by the action of ultraviolet radiation from the sun. Even in this layer, however, the percentage of ozone is only 0.001 by volume. Atmospheric disturbances and downdrafts carry varying amounts of this ozone to the surface of Earth. Human activity adds to ozone in the lower atmosphere, where it becomes a pollutant that can cause extensive crop damage.

The ozone layer became a subject of concern in the early 1970s, when it was found that chemicals known as chlorofluorocarbons (CFCs), or chlorofluoromethanes, were rising into the atmosphere in large quantities because of their use as refrigerants and as propellants in aerosol dispensers. The concern centered on the possibility that these compounds, through the action of sunlight, could chemically attack and destroy stratospheric ozone, which protects Earth`s surface from excessive ultraviolet radiation. As a result, industries in the United States, Europe, and Japan replaced chlorofluorocarbons in all but essential uses.

The atmosphere may be divided into several layers. In the lowest one, the troposphere, the temperature as a rule decreases upward at the rate of 5.5°C per 1,000 m (3°F per 3,000 ft). This is the layer in which most clouds occur. The troposphere extends up to about 16 km (about 10 mi) in tropical regions (to a temperature of about -79°C, or about -110°F) and to about 9.7 km (about 6 mi) in temperate latitudes (to a temperature of about -51°C, or about -60°F). Above the troposphere is the stratosphere. In the lower stratosphere the temperature is practically constant or increases slightly with altitude, especially over tropical regions. Within the ozone layer the temperature rises more rapidly, and the temperature at the upper boundary of the stratosphere, almost 50 km (about 30 mi) above sea level, is about the same as the temperature at the surface of Earth. The layer from 50 to 90 km (30 to 55 mi), called the mesosphere, is characterized by a marked decrease in temperature as the altitude increases.

From investigations of the propagation and reflection of radio waves, it is known that beginning at an altitude of 60 km (40 mi), ultraviolet radiation, X rays, and showers of electrons from the sun ionize several layers of the atmosphere, causing them to conduct electricity; these layers reflect radio waves of certain frequencies back to Earth. Because of the relatively high concentration of ions in the air above 60 km (40 mi), this layer, extending to an altitude of about 1000 km (600 mi), is called the ionosphere. At an altitude of about 90 km (55 mi), temperatures begin to rise. The layer that begins at this altitude is called the thermosphere, because of the high temperatures reached in this layer (about 1200°C, or about 2200°F).
The region beyond the thermosphere is called the exosphere, which extends to about 9,600 km (about 6,000 mi), the outer limit of the atmosphere.

The density of dry air at sea level is about 1/800 the density of water; at higher altitudes it decreases rapidly, being proportional to the pressure and inversely proportional to the temperature. Pressure is measured by a barometer and is expressed in millibars, which are related to the height of a column of mercury that the air pressure will support; 1 millibar equals 0.75 mm (0.03 in) of mercury. Normal atmospheric pressure at sea level is 1,013 millibars, that is, 760 mm (29.92 in) of mercury. At an altitude of 5.6 km (about 3.5 mi) pressure falls to about 507 millibars (about 380 mm/14.96 in of mercury); half of all the air in the atmosphere lies below this level. The pressure is approximately halved for each additional increase of 5.6 km in altitude. At 80 km (50 mi) the pressure is 0.009 millibars (0.0069 mm/0.00027 in of mercury).

The troposphere and most of the stratosphere can be explored directly by means of sounding balloons equipped with instruments to measure the pressure and temperature of the air and with a radio transmitter to send the data to a receiving station at the ground. Rockets carrying radios that transmit meteorological-instrument readings have explored the atmosphere to altitudes above 400 km (250 mi). Study of the form and spectrum of the polar lights gives information to a height possibly as great as 800 km (500 mi).

10. WEATHER

I -INTRODUCTION:

Weather, state of the atmosphere at a particular time and place. The elements of weather include temperature, humidity, cloudiness, precipitation, wind, and pressure. These elements are organized into various weather systems, such as monsoons, areas of high and low pressure, thunderstorms, and tornadoes. All weather systems have well-defined cycles and structural features and are governed by the laws of heat and motion. These conditions are studied in meteorology, the science of weather and weather forecasting.
Weather differs from climate, which is the weather that a particular region experiences over a long period of time. Climate includes the averages and variations of all weather elements.

II -TEMPERATURE:

Temperature is a measure of the degree of hotness of the air. Three different scales are used for measuring temperature. Scientists use the Kelvin, or absolute, scale and the Celsius, or centigrade, scale. Most nations use the Celsius scale, although the United States continues to use the Fahrenheit scale.

Temperature on earth averages 15° C (59° F) at sea level but varies according to latitude, elevation, season, and time of day, ranging from a record high of 58° C (140° F) to a record low of -88° C (-130° F). Temperature is generally highest in the Tropics and lowest near the poles. Each day it is usually warmest during midafternoon and coldest around dawn.

Seasonal variations of temperature are generally more pronounced at higher latitudes. Along the equator, all months are equally warm, but away from the equator, it is generally warmest about a month after the summer solstice (around June 21 in the northern hemisphere and around December 21 in the southern hemisphere) and coldest about a month after the winter solstice (around December 21 in the northern hemisphere and around June 21 in the southern hemisphere). Temperature can change abruptly when fronts (boundaries between two air masses with different temperatures or densities) or thunderstorms pass overhead.

Temperature decreases with increasing elevation at an average rate of about 6.5° C per km (about 19° F per mi). As a result, temperatures in the mountains are generally much lower than at sea level. Temperature continues to decrease throughout the atmosphere's lowest layer, the troposphere, where almost all weather occurs. The troposphere extends to a height of 16 km (10 mi) above sea level over the equator and about 8 km (about 5 mi) above sea level over the poles. Above the troposphere is the stratosphere, where temperature levels off and then begins to increase with height. Almost no weather occurs in the stratosphere.

III -HUMIDITY:

Humidity is a measure of the amount of water vapor in the air. The air's capacity to hold vapor is limited but increases dramatically as the air warms, roughly doubling for each temperature increase of 10° C (18° F). There are several different measures of humidity. The specific humidity is the fraction of the mass of air that consists of water vapor, usually given as parts per thousand. Even the warmest, most humid air seldom has a specific humidity greater than 20 parts per thousand. The most common measure of humidity is the relative humidity, or the amount of vapor in the air divided by the air's vapor-holding capacity at that temperature. If the amount of water vapor in the air remains the same, the relative humidity decreases as the air is heated and increases as the air is cooled. As a result, relative humidity is usually highest around dawn, when the temperature is lowest, and lowest in midafternoon, when the temperature is highest.

IV -CLOUDINESS:

Most clouds and almost all precipitation are produced by the cooling of air as it rises. When air temperature is reduced, excess water vapor in the air condenses into liquid droplets or ice crystals to form clouds or fog. A cloud can take any of several different forms—including cumulus, cirrus, and stratus—reflecting the pattern of air motions that formed it. Fluffy cumulus clouds form from rising masses of air, called thermals. A cumulus cloud often has a flat base, corresponding to the level at which the water vapor first condenses. If a cumulus cloud grows large, it transforms into a cumulonimbus cloud or a thunderstorm. Fibrous cirrus clouds consist of trails of falling ice crystals twisted by the winds. Cirrus clouds usually form high in the troposphere, and their crystals almost never reach the ground. Stratus clouds form when an entire layer of air cools or ascends obliquely. A stratus cloud often extends for hundreds of miles.

Fog is a cloud that touches the ground. In dense fogs, the visibility may drop below 50 m (55 yd). Fog occurs most frequently when the earth's surface is much colder than the air directly above it, such as around dawn and over cold ocean currents. Fog is thickened and acidified when the air is filled with sulfur-laden soot particles produced by the burning of coal. Dense acid fogs that killed thousands of people in London up to 1956 led to legislation that prohibited coal burning in cities.

Optical phenomena, such as rainbows and halos, occur when light shines through cloud particles. Rainbows are seen when sunlight from behind the observer strikes the raindrops falling from cumulonimbus clouds. The raindrops act as tiny prisms, bending and reflecting the different colors of light back to the observer's eye at different angles and creating bands of color. Halos are seen when sunlight or moonlight in front of the observer strikes ice crystals and then passes through high, thin cirrostratus clouds.

V -PRECIPITATION:

Precipitation is produced when the droplets and crystals in clouds grow large enough to fall to the ground. Clouds do not usually produce precipitation until they are more than 1 km (0.6 mi) thick. Precipitation takes a variety of forms, including rain, drizzle, freezing rain, snow, hail, and ice pellets, or sleet. Raindrops have diameters larger than 0.5 mm (0.02 in), whereas drizzle drops are smaller. Few raindrops are larger than about 6 mm (about 0.2 in), because such large drops are unstable and break up easily. Ice pellets are raindrops that have frozen in midair. Freezing rain is rain that freezes on contact with any surface. It often produces a layer of ice that can be very slippery.

Snowflakes are either single ice crystals or clusters of ice crystals. Large snowflakes generally form when the temperature is near 0° C (32° F), because at this temperature the flakes are partly melted and stick together when they collide. Hailstones are balls of ice about 6 to 150 mm (about 0.2 to 6 in) in diameter. They consist of clusters of raindrops that have collided and frozen together. Large hailstones only occur in violent thunderstorms, in which strong updrafts keep the hailstones suspended in the atmosphere long enough to grow large.

Precipitation amounts are usually given in terms of depth. A well-developed winter storm can produce 10 to 30 mm (0.4 to 1.2 in) of rain over a large area in 12 to 24 hours. An intense thunderstorm may produce more than 20 mm (0.8 in) of rain in 10 minutes and cause flash floods (floods in which the water rises suddenly). Hurricanes sometimes produce over 250 mm (10 in) of rain and lead to extensive flooding.

Snow depths are usually much greater than rain depths because of snow's low density. During intense winter storms, more than 250 mm (10 in) of snow may fall in 24 hours, and the snow can be much deeper in places where the wind piles it up in drifts. Extraordinarily deep snows sometimes accumulate on the upwind side of mountain slopes during severe winter storms or on the downwind shores of large lakes during outbreaks of polar air.

VI -WIND:

Wind is the horizontal movement of air. It is named for the direction from which it comes—for example, a north wind comes from the north. In most places near the ground, the wind speed averages from 8 to 24 km/h (from 5 to 15 mph), but it can be much higher during intense storms. Wind speeds in hurricanes and typhoons exceed 120 km/h (75 mph) near the storm's center and may approach 320 km/h (200 mph). The highest wind speeds at the surface of the earth—as high as 480 km/h (300 mph)—occur in tornadoes. Except for these storms, wind speed usually increases with height to the top of the troposphere.

VII -PRESSURE:

Pressure plays a vital role in all weather systems. Pressure is the force of the air on a given surface divided by the area of that surface. In most weather systems the air pressure is equal to the weight of the air column divided by the area of the column. Pressure decreases rapidly with height, halving about every 5.5 km (3.4 mi).

Sea-level pressure varies by only a few percent. Large regions in the atmosphere that have higher pressure than the surroundings are called high-pressure areas. Regions with lower pressure than the surroundings are called low-pressure areas. Most storms occur in low-pressure areas. Rapidly falling pressure usually means a storm is approaching, whereas rapidly rising pressure usually indicates that skies will clear.

VIII -SCALES OF WEATHER:

Weather systems occur on a wide range of scales. Monsoons occur on a global scale and are among the largest weather systems, extending for thousands of miles. Thunderstorms are much smaller, typically 10 to 20 km (6 to 12 mi) across. Tornadoes, which extend from the bases of thunderstorms, range from less than 50 m (55 yd) across to as much as 2 km (1.2 mi) across.
The vertical scale of weather systems is much more limited. Because pressure decreases so rapidly with height and because temperature stops decreasing in the stratosphere, weather systems are confined to the troposphere. Only the tallest thunderstorms reach the stratosphere, which is otherwise almost always clear.

IX -CAUSES OF WEATHER:

All weather is due to heating from the sun. The sun emits energy at an almost constant rate, but a region receives more heat when the sun is higher in the sky and when there are more hours of sunlight in a day. The high sun of the Tropics makes this area much warmer than the poles, and in summer the high sun and long days make the region much warmer than in winter. In the northern hemisphere, the sun climbs high in the sky and the days are long in summer, around July, when the northern end of the earth's axis is tilted toward the sun. At the same time, it is winter in the southern hemisphere. The southern end of the earth's axis is tilted away from the sun, so the sun is low in the sky and the days are short.

The temperature differences produced by inequalities in heating cause differences in air density and pressure that propel the winds. Vertical air motions are propelled by buoyancy: A region of air that is warmer and less dense than the surroundings is buoyant and rises. Air is also forced from regions of higher pressure to regions of lower pressure. Once the air begins moving, it is deflected by the Coriolis force, which results from the earth's rotation. The Coriolis force deflects the wind and all moving objects toward their right in the northern hemisphere and toward their left in the southern hemisphere. It is so gentle that it has little effect on small-scale winds that last less than a few hours, but it has a profound effect on winds that blow for many hours and move over large distances.

X -WEATHER SYSTEMS:

In both hemispheres, the speed of the west wind increases with height up to the top of the troposphere. The core of most rapid winds at the top of the troposphere forms a wavy river of air called the jet stream. Near the ground, where the winds are slowed by friction, the air blows at an acute angle toward areas of low pressure, forming great gyres called cyclones and anticyclones. In the northern hemisphere, the Coriolis force causes air in low-pressure areas to spiral counterclockwise and inward, forming a cyclone, whereas air in high-pressure areas spirals clockwise and outward, forming an anticyclone. In the southern hemisphere, cyclones turn clockwise and anticyclones, counterclockwise.

The air spreading from anticyclones is replaced by sinking air from above. As a result, skies in anticyclones are often fair, and large regions of air called air masses form; these have reasonably uniform temperature and humidity. In cyclones, on the other hand, as air converges to the center, it rises to form extensive clouds and precipitation.

During summer and fall, tropical cyclones, called hurricanes or typhoons, form over warm waters of the oceans in bands parallel to the equator, between about latitude 5° and latitude 30° north and south. Wind speed in hurricanes increases as the air spirals inward. The air either rises in a series of rain bands before reaching the center or proceeds inward and then turns sharply upward in a doughnut-shaped region called the eye wall, where the most intense winds and rain occur. The eye wall surrounds the core, or eye, of the hurricane, which is marked by partly clear skies and gentle winds.

In the middle and high latitudes, polar and tropical air masses are brought together in low-pressure areas called extratropical cyclones, forming narrow zones of sharply changing temperature called fronts. Intense extratropical cyclones can produce blizzard conditions in their northern reaches while at the same time producing warm weather with possible severe thunderstorms and tornadoes in their southern reaches.

Thunderstorms are small, intense convective storms that are produced by buoyant, rapidly rising air. As thunderstorms mature, strong downdrafts of rain- or hail-filled cool air plunge toward the ground, bringing intense showers. However, because thunderstorms are only about 16 km (about 10 mi) wide, they pass over quickly, usually lasting less than an hour. Severe thunderstorms sometimes produce large hail. They may also rotate slowly and spout rapidly rotating tornadoes from their bases.

Most convective weather systems are gentler than thunderstorms. Often, organized circulation cells develop, in which cooler and denser air from the surroundings sinks and blows along the ground to replace the rising heated air. Circulation cells occur on many different scales. On a local scale, along the seashore during sunny spring and summer days, air over the land grows hot while air over the sea remains cool. As the heated air rises, the cooler and denser air from the sea rushes in. This movement of air is popularly called a sea breeze. At night, when the air over the land grows cooler than the air over the sea, the wind reverses and is known as a land breeze.

On a global scale, hot, humid air near the equator rises and is replaced by denser air that sinks in the subtropics and blows back to the equator along the ground. The winds that blow toward the equator are called the trade winds. The trade winds are among the most steady, reliable winds on the earth. They approach the equator obliquely from the northeast and southeast because of the Coriolis force.

The tropical circulation cell is called the Hadley cell. It shifts north and south with the seasons and causes tropical monsoons in India. For example, around July the warm, rising air of the Hadley cell is located over India, and humid winds blow in from the Indian Ocean. Around January the cooler, sinking air of the Hadley cell is located over India, and the winds blow in the opposite direction.

A variable circulation cell called the Walker Circulation exists over the tropical Pacific Ocean. Normally, air rises over the warm waters of the western Pacific Ocean over the Malay Archipelago and sinks over the cold waters in the eastern Pacific Ocean off the coast of Ecuador and Peru. Most years around late December this circulation weakens, and the cold waters off the coast of South America warm up slightly. Because it occurs around Christmas, the phenomenon is called El Niño (The Child). Once every two to five years, the waters of the eastern Pacific Ocean warm profoundly. The Walker Circulation then weakens drastically or even reverses, so that air rises and brings torrential rains to normally dry sections of Ecuador and Peru and hurricanes to Tahiti. On the other side of the Pacific Ocean, air sinks and brings drought to Australia. El Niño can now be predicted with reasonable accuracy several months in advance.

XI -WEATHER FORECASTING:

Since the early 20th century, great strides have been made in weather prediction, largely as a result of computer development but also because of instrumentation such as satellites and radar. Weather data from around the world are collected by the World Meteorological Organization, the National Weather Service, and other agencies and entered into computer models that apply the laws of motion and of the conservation of energy and mass to produce forecasts. In some cases, these forecasts have provided warning of major storms as much as a week in advance. However, because the behavior of weather systems is chaotic, it is impossible to forecast the details of weather more than about two weeks in advance.

Intense small-scale storms, such as thunderstorms and tornadoes, are much more difficult to forecast than are larger weather systems. In areas in which thunderstorms are common, general forecasts can be made several days in advance, but the exact time and location of the storms, as well as of flash floods and tornadoes, can only be forecast about an hour in advance. (For a discussion of weather forecasting methods and technologies, see Meteorology.)

XII -WEATHER MODIFICATION:

Human beings can change weather and climate. Water-droplet clouds with tops colder than about -5° C (about 23° F) can be made to produce rain by seeding them with substances such as silver iodide. Cloud seeding causes ice crystals to form and grow large enough to fall out of a cloud. However, although cloud seeding has been proven effective in individual clouds, its effect over large areas is still unproven.

Weather near the ground is routinely modified for agricultural purposes. For example, soil is darkened to raise its temperature, and fans are turned on during clear, cold nights to stir warmer air down to the ground and help prevent frost damage.

Human activities have also produced inadvertent effects on weather and climate. Adding gases such as carbon dioxide and methane to the atmosphere has increased the greenhouse effect and contributed to global warming by raising the mean temperature of the earth by about 0.5° C (about 0.9° F) since the beginning of the 20th century. More recently, chlorofluorocarbons (CFCs), which are used as refrigerants and in aerosol propellants, have been released into the atmosphere, reducing the amount of ozone worldwide and causing a thinning of the ozone layer over Antarctica each spring (around October). The potential consequences of these changes are vast. Global warming may cause sea level to rise, and the incidence of skin cancer may increase as a result of the reduction of ozone. In an effort to prevent such consequences, production of CFCs has been curtailed and many measures have been suggested to control emission of greenhouse gases, including the development of more efficient engines and the use of alternative energy sources such as solar energy and wind energy

11. CLOUD

I -INTRODUCTION
Cloud, condensed form of atmospheric moisture consisting of small water droplets or tiny ice crystals. Clouds are the principal visible phenomena of the atmosphere. They represent a transitory but vital step in the water cycle, which includes evaporation of moisture from the surface of the earth, carrying of this moisture into higher levels of the atmosphere, condensation of water vapor into cloud masses, and final return of water to the surface as precipitation.

II -FORMATION AND EFFECTS
The formation of clouds caused by cooling of the air results in the condensation of invisible water vapor that produces visible cloud droplets or ice particles. Cloud particles range in size from about 5 to 75 micrometers (0.0005 to 0.008 cm/0.0002 to 0.003 in). The particles are so small that light, vertical currents easily sustain them in the air. The different cloud formations result partly from the temperature at which condensation takes place. When condensation occurs at temperatures below freezing, clouds are usually composed of ice crystals; those that form in warmer air usually consist of water droplets. Occasionally, however, supercooled clouds contain water droplets at subfreezing temperatures. The air motion associated with cloud development also affects formation. Clouds that develop in calm air tend to appear as sheets or stratified formations; those that form under windy conditions or in air with strong vertical currents have a towering appearance.

Clouds perform a very important function in modifying the distribution of solar heat over the earth's surface and within the atmosphere In general, because reflection from the tops of clouds is greater than reflection from the surface of the earth, the amount of solar energy reflected back to space is greater on cloudy days. Although most solar radiation is reflected back by the upper layers of the clouds, some radiation penetrates to the surface of the earth, which absorbs this energy and reradiates it. The lower parts of clouds are opaque to this long-wave earth radiation and reflect it back toward earth.

The result is that the lower atmosphere generally absorbs more radioactive heat energy on a cloudy day because of the presence of this trapped radiation. By contrast, on a clear day more solar radiation is initially absorbed by the surface of the earth, but when reradiated this energy is quickly dissipated because of the absence of clouds. Disregarding related meteorological elements, the atmosphere actually absorbs less radiation on clear days than on cloudy days.

Cloudiness has considerable influence on human activities. Rainfall, which is very important for agricultural activities, has its genesis in the formation of clouds. The marked effect of clouds on visibility at flight levels proved to be a major difficulty during the early days of the airplane, a hazard that was alleviated with the development of instrument flying, which permits the pilot to navigate even in the midst of a thick cloud. The sharp increase in consumption of electricity for lighting during cloudy days represents one of the major scheduling problems faced by the electric-power industry.

The first scientific study of clouds began in 1803, when a method of cloud classification was devised by the British meteorologist Luke Howard. The next development was the publication in 1887 of a classification system that later formed the basis for the noted International Cloud Atlas (1896). This atlas, considerably revised and modified through the years (most recently in 1956), is now used throughout the world.

III -CLASSIFICATION
Clouds are usually divided into four main families on the basis of their height above the ground: high clouds, middle clouds, low clouds, and clouds with vertical development that may extend through all levels. The four main divisions are further subdivided into genera, species, and varieties, which describe in detail the appearance of clouds and the manner in which they are formed. More than 100 different kinds of clouds are distinguishable. Only the primary families and most important genera are described below.

A -High Cloud
These are clouds composed of ice particles, found at average levels of 8 km (5 mi) or more above the earth. The family contains three principal genera. Cirrus clouds are isolated, feathery, and threadlike, often with hooks or tufts, and are arranged in bands. Cirrostratus clouds appear as a fine, whitish veil; they occasionally exhibit a fibrous structure and, when situated between the observer and the moon, produce halo phenomena. Cirrocumulus clouds form small, white, fleecy balls and wisps, arranged in groups or rows. Cirrocumulus and cirrus clouds are popularly described by the phrase ―mackerel scales and mares' tails.‖

B -Middle Clouds
These are clouds composed of water droplets and ranging in altitude from about 3 to 6 km (about 2 to 4 mi) above the earth. Two principal genera are included in the family. Altostratus clouds appear as a thick, gray or bluish veil, through which the sun or moon may be seen only diffusely, as through a frosted glass. Altocumulus clouds have the appearance of dense, fleecy balls or puffs somewhat larger than cirrocumulus. The sun or moon shining through altocumulus clouds may produce a corona, or colored ring, markedly smaller in diameter than a halo.

C -Low Clouds
These clouds, also composed of water droplets, are generally less than 1.6 km (1 mi) high. Three principal forms comprise this group. Stratocumulus clouds consist of large rolls of clouds, soft and gray looking, which frequently cover the entire sky. Because the cloud mass is usually not very thick, blue sky often appears between breaks in the cloud deck. Nimbostratus clouds are thick, dark, and shapeless. They are precipitation clouds from which, as a rule, rain or snow falls. Stratus clouds are sheets of high fog. They appear as flat, white blankets, usually less than 610 m (2000 ft) above the ground. When they are broken up by warm, rising air, the sky beyond usually appears clear and blue.

D -Clouds with Vertical Development
Clouds of this type range in height from less than 1.6 km (1 mi) to more than 13 km (8 mi) above the earth. Two main forms are included in this group. Cumulus clouds are dome-shaped, woolpack clouds most often seen during the middle and latter part of the day, when solar heating produces the vertical air currents necessary for their formation. These clouds usually have flat bases and rounded, cauliflowerlike tops. Cumulonimbus clouds are dark, heavy-looking clouds rising like mountains high into the atmosphere, often showing an anvil-shaped veil of ice clouds, false cirrus, at the top. Popularly known as thunderheads, cumulonimbus clouds are usually accompanied by heavy, abrupt showers.

An anomalous, but exceptionally beautiful, group of clouds contains the nacreous, or mother-of-pearl, clouds, which are 19 to 29 km (12 to 18 mi) high, and the noctilucent clouds, 51 to 56 km (32 to 35 mi) high. These very thin clouds may be seen only between sunset and sunrise and are visible only in high latitudes.

The development of the high-altitude airplane has introduced a species of artificial clouds known as contrails (condensation trails). They are formed from the condensed water vapor ejected as a part of the engine-exhaust gases.

12. RAIN

INTRODUCTION:

Rain, precipitation of liquid drops of water. Raindrops generally have a diameter greater than 0.5 mm (0.02 in). They range in size up to about 3 mm (about 0.13 in) in diameter, and their rate of fall increases, up to 7.6 m (25 ft) per sec with their size. Larger drops tend to be flattened and broken into smaller drops by rapid fall through the air. The precipitation of smaller drops, called drizzle, often severely restricts visibility but usually does not produce significant accumulations of water.

Amount or volume of rainfall is expressed as the depth of water that collects on a flat surface, and is measured in a rain gauge to the nearest 0.25 mm (0.01 in). Rainfall is classified as light if not more than 2.5 mm (0.10 in) per hr, heavy if more than 7.50 mm (more than 0.30 in) per hr, and moderate if between these limits.

PROCESS OF PRECIPITATION:

Air masses acquire moisture on passing over warm bodies of water, or over wet land surfaces. The moisture, or water vapor, is carried upward into the air mass by turbulence and convection (see Heat Transfer). The lifting required cooling and condensing this water vapor results from several processes, and study of these processes provides a key for understanding the distribution of rainfall in various parts of the world.

The phenomenon of lifting associated with the convergence of the trade winds (see Wind), results in a band of copious rains near the equator. This band, called the intertropical convergence zone (ITCZ), moves northward or southward with the seasons. In higher latitudes much of the lifting is associated with moving cyclones (see Cyclone), often taking the form of the ascent of warm moist air, over a mass of colder air, along an interface called a front. Lifting on a smaller scale is associated with convection in air that is heated by a warm underlying surface, giving rise to showers and thunderstorms. The heaviest rainfall over short periods of time usually comes from such storms. Air may also be lifted by being forced to rise over a land barrier, with the result that the exposed windward slopes have enhanced amounts of rain while the sheltered, or lee, slopes have little rain.

AVERAGE RAINFALL:

In the U.S. the heaviest average rainfall amounts, up to 1778 mm (70 in), are experienced in the Southeast, where air masses from the tropical Atlantic and Gulf of Mexico are lifted frequently by cyclones and by convection. Moderate annual accumulations, from 762 to 1270 mm (30 to 50 in), occur throughout the eastern U.S., and are caused by cyclones in winter and convection in summer. The central plains, being farther from sources of moisture, have smaller annual accumulations, 381 to 1016 mm (15 to 40 in), mainly from summer convective storms. The southwestern U.S. is dominated by widespread descent of air in the subtropical Pacific anticyclone; rainfall is light, less than 254 mm (less than 10 in), except in the mountainous regions. The northwestern states are affected by cyclones from the Pacific Ocean, particularly during the winter; but rainfall is moderate, especially on the westward-facing slopes of mountain ranges.

The world's heaviest average rainfall, about 10,922 mm (about 430 in) per year, occurs at Cherrapunji, in northeastern India, where moisture-laden air from the Bay of Bengal is forced to rise over the Khāsi Hills of Assam State. As much as 26,466 mm (1042 in), or 26 m (87 ft), of rain have fallen there in one year. Other extreme rainfall records include nearly 1168 mm (nearly 46 in) of rain in one day during a typhoon at Baguio, Philippines; 304.8 mm (12 in) within one hour during a thunderstorm at Holt, Missouri; and 62.7 mm (2.48in) in over a 5-min period at Portobello, Panama.

ARTIFICIAL PRECIPITATION:

Despite the presence of moisture and lifting, clouds sometimes fail to precipitate rain. This circumstance has stimulated intensive study of precipitation processes, specifically of how single raindrops are produced out of a million or so minute droplets inside clouds. Two precipitation processes are recognized: (1) evaporation of water drops at subfreezing temperatures onto ice crystals that later fall into warmer layers and melt, and (2) the collection of smaller droplets upon larger drops that fall at a higher speed.

Efforts to effect or stimulate these processes artificially have led to extensive weather modification operations within the last 20 years (see Meteorology). These efforts have had only limited success, since most areas with deficient rainfall are dominated by air masses that have either inadequate moisture content or inadequate elevation, or both. Nevertheless, some promising results have been realized and much research is now being conducted in order to develop more effective methods of artificial precipitation

13. ACID RAIN

INTRODUCTION:

Acid Rain, form of air pollution in which airborne acids produced by electric utility plants and other sources fall to Earth in distant regions. The corrosive nature of acid rain causes widespread damage to the environment. The problem begins with the production of sulfur dioxide and nitrogen oxides from the burning of fossil fuels, such as coal, natural gas, and oil, and from certain kinds of manufacturing. Sulfur dioxide and nitrogen oxides react with water and other chemicals in the air to form sulfuric acid, nitric acid, and other pollutants. These acid pollutants reach high into the atmosphere, travel with the wind for hundreds of miles, and eventually return to the ground by way of rain, snow, or fog, and as invisible ―dry‖ forms.

Damage from acid rain has been widespread in eastern North America and throughout Europe, and in Japan, China, and Southeast Asia. Acid rain leaches nutrients from soils, slows the growth of trees, and makes lakes uninhabitable for fish and other wildlife. In cities, acid pollutants corrode almost everything they touch, accelerating natural wear and tear on structures such as buildings and statues. Acids combine with other chemicals to form urban smog, which attacks the lungs, causing illness and premature deaths.

FORMATION OF ACID RAIN:

The process that leads to acid rain begins with the burning of fossil fuels. Burning, or combustion, is a chemical reaction in which oxygen from the air combines with carbon, nitrogen, sulfur, and other elements in the substance being burned. The new compounds formed are gases called oxides. When sulfur and nitrogen are present in the fuel, their reaction with oxygen yields sulfur dioxide and various nitrogen oxide compounds. In the United States, 70 percent of sulfur dioxide pollution comes from power plants, especially those that burn coal. In Canada, industrial activities, including oil refining and metal smelting, account for 61 percent of sulfur dioxide pollution. Nitrogen oxides enter the atmosphere from many sources, with motor vehicles emitting the largest share—43 percent in the United States and 60 percent in Canada.

Once in the atmosphere, sulfur dioxide and nitrogen oxides undergo complex reactions with water vapor and other chemicals to yield sulfuric acid, nitric acid, and other pollutants called nitrates and sulfates. The acid compounds are carried by air currents and the wind, sometimes over long distances. When clouds or fog form in acid-laden air, they too are acidic, and so is the rain or snow that falls from them.

Acid pollutants also occur as dry particles and as gases, which may reach the ground without the help of water. When these —dry‖ acids are washed from ground surfaces by rain, they add to the acids in the rain itself to produce a still more corrosive solution. The combination of acid rain and dry acids is known as acid deposition.

EFFECTS OF ACID RAIN:

The acids in acid rain react chemically with any object they contact. Acids are corrosive chemicals that react with other chemicals by giving up hydrogen atoms. The acidity of a substance comes from the abundance of free hydrogen atoms when the substance is dissolved in water. Acidity is measured using a pH scale with units from 0 to 14. Acidic substances have pH numbers from 1 to 6—the lower the pH number, the stronger, or more corrosive, the substance. Some nonacidic substances, called bases or alkalis, are like acids in reverse—they readily accept the hydrogen atoms that the acids offer. Bases have pH numbers from 8 to 14, with the higher values indicating increased alkalinity. Pure water has a neutral pH of 7—it is not acidic or basic. Rain, snow, or fog with a pH below 5.6 is considered acid rain.

When bases mix with acids, the bases lessen the strength of an acid (see Acids and Bases). This buffering action regularly occurs in nature. Rain, snow, and fog formed in regions free of acid pollutants are slightly acidic, having a pH near 5.6. Alkaline chemicals in the environment, found in rocks, soils, lakes, and streams, regularly neutralize this precipitation. But when precipitation is highly acidic, with a pH below 5.6, naturally occurring acid buffers become depleted over time, and nature's ability to neutralize the acids is impaired. Acid rain has been linked to widespread environmental damage, including soil and plant degradation, depleted life in lakes and streams, and erosion of human-made structures.

A -Soil

In soil, acid rain dissolves and washes away nutrients needed by plants. It can also dissolve toxic substances, such as aluminum and mercury, which are naturally present in some soils, freeing these toxins to pollute water or to poison plants that absorb them. Some soils are quite alkaline and can neutralize acid deposition indefinitely; others, especially thin mountain soils derived from granite or gneiss, buffer acid only briefly.

B -Trees

By removing useful nutrients from the soil, acid rain slows the growth of plants, especially trees. It also attacks trees more directly by eating holes in the waxy coating of leaves and needles, causing brown dead spots. If many such spots form, a tree loses some of its ability to make food through photosynthesis. Also, organisms that cause disease can infect the tree through its injured leaves. Once weakened, trees are more vulnerable to other stresses, such as insect infestations, drought, and cold temperatures.

Spruce and fir forests at higher elevations, where the trees literally touch the acid clouds, seem to be most at risk. Acid rain has been blamed for the decline of spruce forests on the highest ridges of the Appalachian Mountains in the eastern United States. In the Black Forest of southwestern Germany, half of the trees are damaged from acid rain and other forms of pollution.

C -Agriculture

Most farm crops are less affected by acid rain than are forests. The deep soils of many farm regions, such as those in the Midwestern United States, can absorb and neutralize large amounts of acid. Mountain farms are more at risk—the thin soils in these higher elevations cannot neutralize so much acid. Farmers can prevent acid rain damage by monitoring the condition of the soil and, when necessary, adding crushed limestone to the soil to neutralize acid. If excessive amounts of nutrients have been leached out of the soil, farmers can replace them by adding nutrient-rich fertilizer.

D -Surface Waters

Acid rain falls into and drains into streams, lakes, and marshes. Where there is snow cover in winter, local waters grow suddenly more acidic when the snow melts in the spring. Most natural waters are close to chemically neutral, neither acidic nor alkaline: their pH is between 6 and 8. In the northeastern United States and southeastern Canada, the water in some lakes now has a pH value of less than 5 as a result of acid rain. This means they are at least ten times more acidic than they should be. In the Adirondack Mountains of New York State, a quarter of the lakes and ponds are acidic, and many have lost their brook trout and other fish. In the middle Appalachian Mountains, over 1,300 streams are afflicted. All of Norway's major rivers have been damaged by acid rain, severely reducing salmon and trout populations.

E -Plants and Animals

The effects of acid rain on wildlife can be far-reaching. If a population of one plant or animal is adversely affected by acid rain, animals that feed on that organism may also suffer. Ultimately, an entire ecosystem may become endangered. Some species that live in water are very sensitive to acidity, some less so. Freshwater clams and mayfly young, for instance, begin dying when the water pH reaches 6.0. Frogs can generally survive more acidic water, but if their supply of mayflies is destroyed by acid rain, frog populations may also decline. Fish eggs of most species stop hatching at a pH of 5.0. Below a pH of 4.5, water is nearly sterile, unable to support any wildlife.

Land animals dependent on aquatic organisms are also affected. Scientists have found that populations of snails living in or near water polluted by acid rain are declining in some regions. In The Netherlands songbirds are finding fewer snails to eat. The eggs these birds lay have weakened shells because the birds are receiving less calcium from snail shells.

F -Human-Made Structures

Acid rain and the dry deposition of acidic particles damage buildings, statues, automobiles, and other structures made of stone, metal, or any other material exposed to weather for long periods. The corrosive damage can be expensive and, in cities with very historic buildings, tragic. Both the Parthenon in Athens, Greece, and the Taj Mahal in Agra, India, are deteriorating due to acid pollution.

G -Human Health

The acidification of surface waters causes little direct harm to people. It is safe to swim in even the most acidified lakes. However, toxic substances leached from soil can pollute local water supplies. In Sweden, as many as 10,000 lakes have been polluted by mercury released from soils damaged by acid rain, and residents have been warned to avoid eating fish caught in these lakes. In the air, acids join with other chemicals to produce urban smog, which can irritate the lungs and make breathing difficult, especially for people who already have asthma, bronchitis, or other respiratory diseases. Solid particles of sulfates, a class of minerals derived from sulfur dioxide, are thought to be especially damaging to the lungs.

H -Acid Rain and Global Warming

Acid pollution has one surprising effect that may be beneficial. Sulfates in the upper atmosphere reflect some sunlight out into space, and thus tend to slow down global warming. Scientists believe that acid pollution may have delayed the onset of warming by several decades in the middle of the 20th century.

EFFORTS TO CONTROL ACID RAIN:

Acid rain can best be curtailed by reducing the amount of sulfur dioxide and nitrogen oxides released by power plants, motorized vehicles, and factories. The simplest way to cut these emissions is to use less energy from fossil fuels. Individuals can help. Every time a consumer buys an energy-efficient appliance, adds insulation to a house, or takes a bus to work, he or she conserves energy and, as a result, fights acid rain.

Another way to cut emissions of sulfur dioxide and nitrogen oxides is by switching to cleaner-burning fuels. For instance, coal can be high or low in sulfur, and some coal contains sulfur in a form that can be washed out easily before burning. By using more of the low-sulfur or cleanable types of coal, electric utility companies and other industries can pollute less. The gasoline and diesel oil that run most motor vehicles can also be formulated to burn more cleanly, producing less nitrogen oxide pollution. Clean-burning fuels such as natural gas are being used increasingly in vehicles. Natural gas contains almost no sulfur and produces very low nitrogen oxides. Unfortunately, natural gas and the less-polluting coals tend to be more expensive, placing them out of the reach of nations that are struggling economically.

Pollution can also be reduced at the moment the fuel is burned. Several new kinds of burners and boilers alter the burning process to produce less nitrogen oxides and more free nitrogen, which is harmless. Limestone or sandstone added to the combustion chamber can capture some of the sulfur released by burning coal.

Once sulfur dioxide and oxides of nitrogen have been formed, there is one more chance to keep them out of the atmosphere. In smokestacks, devices called scrubbers spray a mixture of water and powdered limestone into the waste gases (flue gases), recapturing the sulfur. Pollutants can also be removed by catalytic converters. In a converter, waste gases pass over small beads coated with metals. These metals promote chemical reactions that change harmful substances to less harmful ones. In the United States and Canada, these devices are required in cars, but they are not often used in smokestacks.

Once acid rain has occurred, a few techniques can limit environmental damage. In a process known as liming, powdered limestone can be added to water or soil to neutralize the acid dropping from the sky. In Norway and Sweden, nations much afflicted with acid rain, lakes are commonly treated this way. Rural water companies may need to lime their reservoirs so that acid does not eat away water pipes. In cities, exposed surfaces vulnerable to acid rain destruction can be coated with acid-resistant paints. Delicate objects like statues can be sheltered indoors in climate-controlled rooms.

Cleaning up sulfur dioxide and nitrogen oxides will reduce not only acid rain but also smog, which will make the air look clearer. Based on a study of the value that visitors to national parks place on clear scenic vistas, the U.S. Environmental Protection Agency thinks that improving the vistas in eastern national parks alone will be worth $1 billion in tourist revenue a year.

A -National Legislation

In the United States, legislative efforts to control sulfur dioxide and nitrogen oxides began with passage of the Clean Air Act of 1970. This act established emissions standards for pollutants from automobiles and industry. In 1990 Congress approved a set of amendments to the act that impose stricter limits on pollution emissions, particularly pollutants that cause acid rain. These amendments aim to cut the national output of sulfur dioxide from 23.5 million tons to 16 million tons by the year 2010. Although no national target is set for nitrogen oxides, the amendments require that power plants, which emit about one-third of all nitrogen oxides released to the atmosphere, reduce their emissions from 7.5 million tons to 5 million tons by 2010. These rules were applied first to selected large power plants in Eastern and Midwestern states. In the year 2000, smaller, cleaner power plants across the country came under the law.

These 1990 amendments include a novel provision for sulfur dioxide control. Each year the government gives companies permits to release a specified number of tons of sulfur dioxide. Polluters are allowed to buy and sell their emissions permits. For instance, a company can choose to reduce its sulfur dioxide emissions more than the law requires and sell its unused pollution emission allowance to another company that is further from meeting emission goals; the buyer may then pollute above the limit for a certain time. Unused pollution rights can also be "banked" and kept for later use. It is hoped that this flexible market system will clean up emissions more quickly and cheaply than a set of rigid rules.

Legislation enacted in Canada restricts the annual amount of sulfur dioxide emissions to 2.3 million tons in all of Canada's seven easternmost provinces, where acid rain causes the most damage. A national cap for sulfur dioxide emissions has been set at 3.2 million tons per year. Legislation is currently being developed to enforce stricter pollution emissions by 2010.
Norwegian law sets the goal of reducing sulfur dioxide emission to 76 percent of 1980 levels and nitrogen oxides emissions to 70 percent of the 1986 levels. To encourage cleanup, Norway collects a hefty tax from industries that emit acid pollutants. In some cases these taxes make it more expensive to emit acid pollutants than to reduce emissions.

B -International Agreements
Acid rain typically crosses national borders, making pollution control an international issue. Canada receives much of its acid pollution from the United States—by some estimates as much as 50 percent. Norway and Sweden receive acid pollutants from Britain, Germany, Poland, and Russia. The majority of acid pollution in Japan comes from China. Debates about responsibilities and cleanup costs for acid pollutants led to international cooperation. In 1988, as part of the Long-Range Transboundary Air Pollution Agreement sponsored by the United Nations, the United States and 24 other nations ratified a protocol promising to hold yearly nitrogen oxide emissions at or below 1987 levels. In 1991 the United States and Canada signed an Air Quality Agreement setting national limits on annual sulfur dioxide emissions from power plants and factories. In 1994 in Oslo, Norway, 12 European nations agreed to reduce sulfur dioxide emissions by as much as 87 percent by 2010.

Legislative actions to prevent acid rain have results. The targets established in laws and treaties are being met, usually ahead of schedule. Sulfur emissions in Europe decreased by 40 percent from 1980 to 1994. In Norway sulfur dioxide emissions fell by 75 percent during the same period. Since 1980 annual sulfur dioxide emissions in the United States have dropped from 26 million tons to 18.3 million tons. Canada reports sulfur dioxide emissions have been reduced to 2.6 million tons, 18 percent below the proposed limit of 3.2 million tons.

Monitoring stations in several nations report that precipitation is actually becoming less acidic. In Europe, lakes and streams are now growing less acid. However, this does not seem to be the case in the United States and Canada. The reasons are not completely understood, but apparently, controls reducing nitrogen oxide emissions only began recently and their effects have yet to make a mark. In addition, soils in some areas have absorbed so much acid that they contain no more neutralizing alkaline chemicals. The weathering of rock will gradually replace the missing alkaline chemicals, but scientists fear that improvement will be very slow unless pollution controls are made even stricter.

14. MINERALS

A mineral is an inorganic substance formed naturally. They are infact building blocks from which rocks are made and they may form crystals.

Minerals are produced by the physical force of the earth. They are the basic units of rocks. Every rock can be considered as the accumulation of minerals. Geologists classify rock principally according to the way which they are formed, not according to their composition as most rocks consist of two or more than two different minerals.

As rocks are formed by minerals, therefore there are three kinds of rocks igneous, sedimentary and metamorphic.

The igneous rocks are formed when the molten magma cools and solidifies.
The sedimentary rocks originate from the deposits of material worn away from the plants and animals and also by the pre existing rocks.
Metamorphic rocks are formed by the alteration of pre existing rocks by great heat or pressure.

The abundant element in these rocks is oxygen and silicon. Because of this the mineral silica is very often found in the rocks. In the rocks the silica is combined with other elements such as aluminium, calcium, iron, magnesium, pottasium, or sodium

15. ROCK (MINERAL)

I -INTRODUCTION
Rock (mineral), naturally occurring solid material consisting of one or more minerals. Minerals are solid, naturally occurring chemical elements or compounds that are homogenous, meaning they have a definite chemical composition and a very regular arrangement of atoms. Rocks are everywhere, in the ground, forming mountains, and at the bottom of the oceans. Earth's outer layer, or crust, is made mostly of rock. Some common rocks include granite and basalt.

II -TYPES OF ROCKS
Rocks are divided into three main types, based on the ways in which they form. Igneous rocks are made of old rocks that have melted within the earth to form molten material called magma. Magma cools and solidifies to become igneous rocks. Sedimentary rocks form as layers of material settle onto each other, press together, and harden. Metamorphic rocks are created when existing rocks are exposed to high temperatures and pressures, and the rock material is changed, or metamorphosed, while solid.

A -Igneous Rock
Igneous rocks are rocks formed from a molten or partly molten material called magma. Magma forms deep underground when rock that was once solid melts. Overlying rock presses down on the magma, and the less dense magma rises through cracks in the rock. As magma moves upward, it cools and solidifies. Magma that solidifies underground usually cools slowly, allowing large crystals to form. Magma that reaches Earth's surface is called lava. Lava loses heat to the atmosphere or ocean very quickly and therefore solidifies very rapidly, forming very small crystals or glass. When lava erupts at the surface again and again, it can form

mountains called volcanoes.

Igneous rocks commonly contain the minerals feldspar, quartz, mica, pyroxene, amphibole, and olivine. Igneous rocks are named according to which minerals they contain. Rocks rich in feldspar and quartz are called felsic; rocks rich in pyroxene, amphibole, and olivine, which all contain magnesium and iron, are called mafic. Common and important igneous rocks are granite, rhyolite, gabbro, and basalt. Granite and rhyolite are felsic; gabbro and basalt are mafic. Granite has large crystals of quartz and feldspar. Rhyolite is the small-grained equivalent of granite. Gabbro has large crystals of pyroxene and olivine. Basalt is the most common volcanic rock.

B -Sedimentary Rock

Sedimentary rock forms when loose sediment, or rock fragments, hardens. Geologists place sedimentary rocks into three broad categories: (1) clastic rocks, which form from clasts, or broken fragments, of pre-existing rocks and minerals; (2) chemical rocks, which form when minerals precipitate, or solidify, from a solution, usually seawater or lake water; and (3) organic rocks, which form from accumulations of animal and plant remains. It is common for sedimentary rocks to contain all three types of sediment. Most fossils are found in sedimentary rocks because the processes that form igneous and metamorphic rocks prevent fossilization or would likely destroy fossils.

The most common types of clastic rocks are sandstone and shale (also known as mudrock). Sandstone is made from sand, and shale is made from mud. Sand particles have diameters in the range 2.00 to 0.06 mm (0.08 to 0.002 in), while mud particles are smaller than 0.06 mm (0.002 in). Sand and mud form when physical or chemical processes break down and destroy existing rocks. The sand and mud are carried by wind, rivers, ocean currents, and glaciers, which deposit the sediment when the wind or water slows down or where the glacier ends. Sand usually forms dunes in deserts, or sandbars, riverbeds, beaches, and nearshore marine deposits. Mud particles are smaller than sand particles, so they tend to stay in the wind or water longer and are deposited only in very still environments, such as lake beds and the ocean floor.

Sedimentary rock forms when layers of sand and mud accumulate. As the sediment accumulates, the weight of the layers of sediment presses down and compacts the layers underneath. The sediments become cemented together into a hard rock when minerals (most commonly quartz or calcite) precipitate, or harden, from water in the spaces between grains of sediment, binding the grains together. Sediment is usually deposited in layers, and compaction and cementation preserve these layers, called beds, in the resulting sedimentary rock.

The most common types of chemical rocks are called evaporites because they form by evaporation of seawater or lake water. The elements dissolved in the water crystallize to form minerals such as gypsum and halite. Gypsum is used to manufacture plaster and wallboard; halite is used as table salt.

The most common organic rock is limestone. Many marine animals, such as corals and shellfish, have skeletons or shells made of calcium carbonate ($CaCO_3$). When these animals die, their skeletons sink to the seafloor and accumulate to form large beds of calcium carbonate. As more and more layers form, their weight compresses and cements the layers at the bottom, forming limestone. Details of the skeletons and shells are often preserved in the limestone as fossils.

Coal is another common organic rock. Coal comes from the carbon compounds of plants growing in swampy environments. Plant material falling into the muck at the bottom of the swamp is protected from decay. Burial and compaction of the accumulating plant material can produce coal, an important fuel in many parts of the world. Coal deposits frequently contain plant fossils.

C -Metamorphic Rock

Metamorphic rock forms when pre-existing rock undergoes mineralogical and structural changes resulting from high temperatures and pressures. These changes occur in the rock while it remains solid (without melting).

The changes can occur while the rock is still solid because each mineral is stable only over a specific range of temperature and pressure. If a mineral is heated or compressed beyond its stability range, it breaks down and forms another mineral. For example, quartz is stable at room temperature and at pressures up to 1.9 gigapascals (corresponding to the pressure found about 65 km [about 40 mi] underground). At pressures above 1.9 gigapascals, quartz breaks down and forms the mineral coesite, in which the silicon and oxygen atoms are packed more closely together.

In the same way, combinations of minerals are stable over specific ranges of temperature and pressure. At temperatures and pressures outside the specific ranges, the minerals react to form different combinations of minerals. Such combinations of minerals are called mineral assemblages.

In a metamorphic rock, one mineral assemblage changes to another when its atoms move about in the solid state and recombine to form new minerals. This change from one mineral assemblage to another is called metamorphism. As temperature and pressure increase, the rock gains energy, which fuels the chemical reactions that cause metamorphism. As temperature and pressure decrease, the rock cools; often, it does not have enough energy to change back to a low-temperature and low-pressure mineral assemblage. In a sense, the rock is stuck in a state that is characteristic of its earlier high-temperature and high-pressure environment. Thus, metamorphic rocks carry with them information about the history of temperatures and pressures to which they were subjected.

The size, shape, and distribution of mineral grains in a rock are called the texture of the rock. Many metamorphic rocks are named for their main texture. Textures give important clues as to how the rock formed. As the pressure and temperature that form a metamorphic rock increase, the size of the mineral grains usually increases. When the pressure is equal in all directions, mineral grains form in random orientations and point in all directions. When the pressure is stronger in one direction than another, minerals tend to align themselves in particular directions. In particular, thin plate-shaped minerals, such as mica, align perpendicular to the direction of maximum pressure, giving rise to a layering in the rock that is known as foliation. Compositional layering, or bands of different minerals, can also occur and cause foliation. At low pressure, foliation forms fine, thin layers, as in the rock slate. At medium pressure, foliation becomes coarser, forming schist. At high pressure, foliation is very coarse, forming gneiss. Commonly, the layering is folded in complex, wavy patterns from the pressure.

III -THE ROCK CYCLE
The rock cycle describes how rocks change, or evolve, from one type to another. For example, any type of rock (igneous, sedimentary, or metamorphic) can become a new sedimentary rock if its eroded sediments are deposited, compacted, and cemented. Similarly, any type of rock can become metamorphic if it is buried moderately deep. If the temperature and pressure become sufficiently high, the rock can melt to form magma and a new igneous rock.

16. MINERAL DEPOSIT

I -INTRODUCTION
Mineral Deposit, concentrated, natural occurrence of one or more minerals. Mineral deposits can form within any kind of rock and consist of any type of mineral. They are valuable economically because they contain high concentrations of metallic and nonmetallic elements or other valuable materials that are essential to an industrial society.

The concentration of a mineral in a mineral deposit is critically important in determining whether it can be mined profitably. For the mining of metals, concentration in a mineral deposit is measured two ways. The grade depends on the percentage by weight of a metal in a mineral deposit. This percentage is measured by dividing the weight of the metal by the weight of the rock. The concentration factor (also called enrichment factor) is the number of times more abundant a metal is in a mineral deposit than it is in average crustal rock. The concentration factor is measured by dividing a mineral deposit‟s grade by the average grade of crustal rocks for that metal. A concentration factor of ten, for example, means that a metal is ten times more abundant in a particular deposit than in the earth‟s crust.

If a metal is to be mined profitably, it must have attained a minimum concentration factor—otherwise, the amount of that metal acquired will be too small to pay for the mining process. Minimum concentration factors vary from one metal to the next. Iron, which is relatively abundant in the earth‟s crust, typically requires a concentration factor of between 5 and 10. Gold and silver, however, require concentration factors in excess of 2,000. The term ore describes rock that contains high enough concentrations of a metal to be mined profitably.

The accessibility of a mineral deposit also plays an important role in determining the cost-effectiveness of mining. In general, deposits that reside deeper in the crust are more difficult and more expensive to mine. Consequently, the minimum required concentration factor increases with the difficulty of extraction.

II -PROCESSES OF SEGREGATION
Geological processes, such as melting and crystallizing of igneous rocks as well as erosion and deposition, sometimes separate and concentrate minerals. At other times, these processes mix and dilute them. Any process that separates and concentrates minerals is called a process of segregation.

A -Magmatic Processes
During cooling and crystallization of a magma, minerals with a high temperature of crystallization form early and may settle to the floor of the magma chamber. These early-formed minerals, such as pyroxene or olivine, tend to be relatively rich in iron and magnesium and poor in silicon and oxygen when compared to the entire magma. They also typically contain no potassium or aluminum. Consequently, minerals with lower temperatures of crystallization that form later tend to be relatively rich in potassium, aluminum, silicon, and oxygen, but poor in iron and magnesium. This process, called fractional crystallization, segregates minerals.

Fractional crystallization can lead to valuable mineral deposits because many rare and valuable elements form mineral crystals either early or late in the crystallization process. For example, when magmas have compositions with abundant chromium, the mineral chromite crystallizes early and can form deposits on the floor of the magma chamber. Extensive chromite deposits are mined in the Bushveld Complex of South Africa and in the Stillwater Complex of Montana, United States. In other magmas, the latest-forming mineral crystals may contain a variety of rare elements such as beryllium, lithium, boron, molybdenum, and uranium. These deposits are called pegmatites. Numerous well-known pegmatites are scattered throughout the western United States.

B -Hydrothermal Processes
Hydrothermal processes involve the transportation of elements dissolved in hot water and the subsequent precipitation, or crystallization, of minerals when the water cools. In some cases, the elements precipitate in their native states, such as pure gold or copper. More often, however, they precipitate as sulfide minerals, including pyrite (iron sulfide), galena (lead sulfide), sphalerite (zinc sulfide), cinnabar (mercury sulfide), and chalcopyrite (copper sulfide). Hydrothermal processes are particularly effective at segregating minerals because the fluids typically contain only a small variety of dissolved elements. Hydrothermal processes are responsible for most of the world‟s metallic mineral deposits such as gold, silver, lead, and copper.

Hydrothermal fluids originate in several different ways. Some originate from magmas that have water dissolved in them. As the magma cools and crystallizes, the water is excluded from the growing crystals and separates from the magma. Such fluids will be very hot and rich with elements dissolved from the magma. Other sources of hydrothermal fluids include circulating groundwater that comes into contact with hot rock, or seawater circulating through seafloor sediments that interacts with newly created volcanic rock on the ocean floor. These fluids typically migrate away from their heat sources along fractures and cool. This cooling causes some minerals to precipitate.

When minerals form a precipitate within open fractures, the resulting deposit is called a vein. During the late 19th and early 20th centuries, miners exploited veins of highly concentrated gold throughout the western United States. Two well-known examples are Cripple Creek in Colorado, and Bullfrog in Nevada. Besides cooling, other causes of precipitation include sudden decreases in pressure or reactions with the surrounding rock. When precipitation occurs at the earth's surface, the minerals form hot springs deposits, such as the deposits at Yellowstone National Park.
Precipitation of minerals can also occur in microscopic networks of fractures or pore spaces to form mineral deposits that are disseminated, or spread widely, throughout the rock. Disseminated deposits typically display much lower concentration factors than vein deposits. Some are so extensive, however, that their huge volumes make up for the low concentrations. Many of the copper mines in Arizona and Utah, and many of the gold mines in Nevada, are in disseminated deposits.

C -Evaporation Processes
When water containing dissolved minerals evaporates, the minerals will precipitate. Deposits of minerals formed in this way are called evaporites. Evaporite deposits can form on land in enclosed arid basins. Incoming water cannot exit except by evaporation. Because the incoming water also carries dissolved minerals, the basin continually receives additional minerals, and the resulting deposit can be quite thick. Land-based evaporites currently are forming in desert lakes in the American states of California, Nevada, and Utah, and in the Dead Sea between Israel and Jordan.

Evaporite deposits also form in tropical seas or bays connected to the open ocean through narrow passages. Seawater flows through the narrow passages to replace water lost through evaporation. Because the incoming water is salty, the basin continually receives additional sea salts. If the concentration of salts is high enough, the minerals will precipitate. If the conditions persist for a long time, the resultant deposits can be very thick. In the western United States, a thick layer of marine evaporites formed more than 200 million years ago during the Permian Period.

Some examples of common evaporite minerals are halite (sodium chloride), gypsum (calcium sulfate), and borax (sodium borate). Many evaporite deposits are mined for use in table salt, fertilizers, wallboard, plaster, detergents, and fluxes.

D -Residues of Weathering Process
Chemical weathering causes minerals to decompose into clays and other materials. This weathering typically leads to the removal of all material that does not resist weathering. In regions of especially intense weathering, such as the tropics, virtually everything except oxides of aluminum and iron becomes weathered and is eventually removed. Through this process of weathering and removal of the nonresistant material, aluminum and iron oxides form a concentrated residue. These residues, if extensive enough, can be mined for aluminum and iron.

Bauxite is a rock made from aluminum oxide residues and is the principal ore of aluminum. The world's leading producers of bauxite, the countries Surinam, Jamaica, and Guyana, are all located in the tropics. Commercial bauxite deposits that occur outside of the tropics, such as in the United States, the former Soviet Union, and China, indicate that those regions once experienced tropical weathering conditions.

E -Depositional Processes
Some mineral deposits form in river beds because running water tends to segregate dense minerals. Rivers deposit grains that are either larger or denser first, and then carry grains that are either smaller or lighter farther downriver. Relatively dense minerals or metals, such as cassiterite (a source of tin), diamond, or gold, erode from their sources and get deposited with the heavier, coarse grains. The sites of deposition are most frequently the gravel or sandbars that form on the inside bends of meandering rivers. Mineable deposits of these materials are called placer deposits.

Placer mining has provided humankind with more than half of its gold. Well-known placer deposits include gravels formed about 40 million years ago during the Eocene Epoch in California, the discovery of which helped fuel the 1849 California Gold Rush. Much of this placer gold originally eroded from hydrothermal vein deposits of gold associated with igneous intrusions in western Nevada. Precambrian deposits in South Africa, formed more than 500 million years ago, are the largest known placer gold deposits in the world

17. LAVA

I -INTRODUCTION:

Lava, molten or partially molten rock that erupts at the earth's surface. When lava comes to the surface, it is red-hot, reaching temperatures as high as 1200° C (2200° F). Some lava can be as thick and viscous as toothpaste, while other lava can be as thin and fluid as warm syrup and flow rapidly down the sides of a volcano. Molten rock that has not yet erupted is called magma.
Once lava hardens it forms igneous rock. Volcanoes build up where lava erupts from a central vent. Flood basalt forms where lava erupts from huge fissures. The eruption of lava is the principal mechanism whereby new crust is produced (see Plate Tectonics). Since lava is generated at depth, its chemical and physical characteristics provide indirect information about the chemical composition and physical properties of the rocks 50 to 150 km (30 to 90 mi) below the surface.

II -TYPES OF LAVA:

Most lava, on cooling, forms silicate rocks—rocks that contain silicon and oxygen. Lava is classified according to which silicate rocks it forms: basalt, rhyolite, or andesite. Basaltic lava is dark in color and rich in magnesium and iron, but poor in silicon. Rhyolitic lava is light colored and poor in magnesium and iron, but rich in silicon. Andesitic lava is intermediate in composition between basaltic and rhyolitic lava. While color is often sufficient to classify lava informally, formal identification requires chemical analysis in a laboratory. If silica (silicon dioxide) makes up more than 65 percent of the weight of the lava, then the lava is rhyolitic. If the silica content is between 65 percent and 50 percent by weight, then the lava is andesitic. If the silica content is less than 50 percent by weight, then the lava is basaltic.

Many other physical properties, in addition to color, follow the distinctions between basaltic, andesitic, and rhyolitic lava. For example, basaltic lava has a low viscosity, meaning it is thin and runny. Basaltic lava flows easily and spreads out. Rhyolitic lava has a high viscosity and oozes slowly like toothpaste. The viscosity of andesitic lava is intermediate between basaltic and rhyolitic lava. Similarly, basaltic lava tends to erupt at higher temperatures, typically around 1000° to 1200° C (1800° to 2200° F), while rhyolitic lava tends to erupt at temperatures of 800° to 1000° C (1500° to 1800° F). Dissolved gases make up between 1 percent and 9 percent of magma. These gases come out of solution and form gas bubbles as the magma nears the surface. Rhyolitic lava tends to contain the most gas and basaltic lava tends to contain the least.

III -ERUPTIVE STYLES:

Lava can erupt in several different ways depending on the viscosity of the lava and the pressure from the overlaying rock. When lava erupts out of a vent or large crack, it may pour like water out of a large pipe. The lava flows downhill like a river and can also form large lava lakes. The rivers and lakes of lava are called lava flows. Other times, the pressure exerted by gas bubbles in the lava is so high that it shatters the overlying rock and shoots lava and rock fragments high into the air with explosive force. The fragments of hot rock and lava shot into the air are called pyroclasts (Greek pyro, ―fire‖; and klastos, ―fragment‖). At other times, the pressure may be so high that the volcano itself is destroyed in a cataclysmic explosion.

A -Lava Flows
When lava flows out of a central vent, it forms a volcano. Basaltic lava is thin and fluid so it quickly spreads out and forms gently sloping volcanoes with slopes of about 5°. The flattest slopes are nearest the top vent, where the lava is hottest and most fluid. These volcanoes are called shield volcanoes because from a distance, they look like giant shields lying on the ground. Mauna Kea and Mauna Loa, on the island of Hawaii, are classic examples of shield volcanoes. Andesitic lava is more viscous and does not travel as far, so it forms steeper volcanoes. Rhyolitic lava is so viscous it does not flow away from the vent. Instead, it forms a cap or dome over the vent.

Sometimes, huge amounts of basaltic lava flow from long cracks or fissures in the earth. These basaltic lava flows, known as flood basalts, can cover more than 100,000 sq km (40,000 sq mi) to a depth of more than 100 m (300 ft). The Columbia River plateau in the states of Washington, Oregon, and Idaho was formed by repeated fissure eruptions. The accumulated basalt deposits are more than 4000 m (13,000 ft) thick in places and cover more than 200,000 sq km (80,000 sq mi). The Parana of Brazil and Paraguay covers an area four times as large. Flood basalts occur on every continent. When basaltic lava cools, it shrinks. In thick sheets of basaltic lava, this shrinking can produce shrinkage cracks that often occur in a hexagonal pattern and create hexagonal columns of rock, a process known as columnar jointing.

Two well-known examples of columnar jointing are the Giant‗s Causeway on the coast of Northern Ireland and Devil‗s Tower in northeastern Wyoming.

Basaltic lava flows and rocks are classified according to their texture. Pahoehoe flows have smooth, ropy-looking surfaces. They form when the semi cooled; semi hard surface of a lava flow is twisted and wrinkled by the flow of hot fluid lava beneath it. Fluid lava can drain away from beneath hardened pahoehoe surfaces to form empty lava tubes and lava caves. Other basaltic lava flows, known as aa flows, have the appearance of jagged rubble. Very fast-cooling lava can form volcanic glass, such as obsidian.

Vesicular basalt, or scoria, is a solidified froth formed when bubbles of gas trapped in the basaltic lava rise to the surface and cool. Some gas-rich andesitic or rhyolitic lava produces rock, called pumice that has so many gas bubbles that it will float in water.

Pillow lava is made up of interconnected pillow-shaped and pillow-sized blocks of basalt. It forms when lava erupts underwater. The surface of the lava solidifies rapidly on contact with the water, forming a pillow-shaped object. Pressure of erupting lava beneath the pillow causes the lava to break through the surface and flow out into the water, forming another pillow. Repetition of this process gives rise to piles of pillows. Pillow basalts cover much of the ocean floor.

B -Pyroclastic Eruptions
Pyroclasts are fragments of hot lava or rock shot into the air when gas-rich lava erupts. Gases easily dissolve in liquids under pressure and come out of solution when the pressure is released. Magma deep underground is under many tons of pressure from the overlying rock. As the magma rises, the pressure from the overlying rocks drops because less weight is pressing down on the magma. Just as the rapid release of bubbles can force a fountain of soda to be ejected from a shaken soda bottle, the rapid release of gas can propel the explosive release of lava.

Pyroclasts come in a wide range of sizes, shapes, and textures. Pieces smaller than peas are called ash. Cinders are pea sized to walnut sized, and anything larger are lava bombs.

Cinders and bombs tend to fall to earth fairly close to where they are ejected, but in very strong eruptions they can travel farther. Lava bombs as large as 100 tons have been found 10 km (6 mi) from the volcano that ejected them. When cinders and bombs accumulate around a volcanic vent, they form a cinder cone. Although the fragments of lava cool rapidly during their brief flight through the air, they are usually still hot and sticky when they land. The sticky cinders weld together to form a rock called tuff. Ash, because it is so much smaller than cinders, can stay suspended in the air for hours or weeks and travel great distances. The ash from the 1980 eruption of Mount Saint Helens in the state of Washington circled the earth twice.

Many volcanoes have both lava eruptions and pyroclastic eruptions. The resulting volcano is composed of alternating layers of lava and pyroclastic material. These volcanoes are called composite volcanoes or stratovolcanoes. With slopes of 15° to 20°, they are steeper than the gently sloped shield volcanoes. Many stratovolcanoes, such as the picturesque Mount Fuji in Japan, have convex slopes that get steeper closer to the top.

Pyroclastic materials that accumulate on the steep upper slopes of stratovolcanoes often slide down the mountain in huge landslides. If the volcano is still erupting and the loose pyroclastic material is still hot, the resulting slide is called a pyroclastic flow or nuée ardente (French for "glowing cloud"). The flow contains trapped hot gases that suspend the ash and cinders, enabling the flow to travel at great speed. Such flows have temperatures of 800° C (1500° F) and often travel in excess of 150 km/h (100 mph). One such pyroclastic flow killed 30,000 people in the city of Saint-Pierre on the Caribbean island of Martinique in 1902. Only one person in the whole town survived. He was in a basement jail cell.

Loose accumulations of pyroclastic material on steep slopes pose a danger long after the eruption is over. Heavy rains or melting snows can turn the material into mud and set off a catastrophic mudflow called a lahar. In 1985 a small pyroclastic eruption on Nevado del Ruiz, a volcano in Colombia, melted snowfields near the summit. The melted snow, mixed with new and old pyroclastic material, rushed down the mountain as a wall of mud 40 m (140 ft) tall. One hour later, it smashed into the town of Armero 55 km (35 mi) away, killing 23,000 people.

C -Explosive Eruptions
Rhyolitic lava, because it is so viscous, and because it contains so much gas, is prone to cataclysmic eruption. The small amount of lava that does emerge from the vent is too thick to spread. Instead it forms a dome that often caps the vent and prevents the further release of lava or gas. Gas and pressure can build up inside the volcano until the mountaintop blows apart. Such an eruption occurred on Mount Saint Helens in 1980, blowing off the top 400 m (1,300 ft) of the mountain.

Other catastrophic eruptions, called phreatic explosions, occur when rising magma reaches underground water. The water rapidly turns to steam which powers the explosion. One of the most destructive phreatic explosions of recorded history was the 1883 explosion of Krakatau, in the strait between the Indonesian islands of Java and Sumatra. It destroyed most of the island of Krakatau. The island was uninhabited, so no one died in the actual explosion. However, the explosion caused tsunamis (giant ocean waves) that reached an estimated height of 30 m (100 ft) and hit the nearby islands of Sumatra and Java, destroying 295 coastal towns and killing about 34,000 people. The noise from the explosion was heard nearly 2,000 km (1,200 mi) away in Australia

18. SOLAR AND LUNAR ECLIPSES

ECLIPSE:
An eclipse is a New or Full Moon that occurs near the Moon's Nodes. Because the Moon and Sun are so close to the Nodes, they are aligned perfectly enough with the Earth to cast a shadow.

SOLAR ECLIPSE:

The solar eclipse takes place when the light of the sun is partially or totally cut off from the earth by the moon which comes in between both these celestial bodies.i.e. sun and the earth.

However the solar eclipse occurs only at new moon. As moon is of the similar size to that of the sun, therefore when it passes directly between the earth and the sun it obscures it completely.

There are three types of eclipse.
1. Total
2.Partial
3.annular

the total eclipse is possible when the apparent sizes of both sun and moon are equal to each other as the moon can completely obscure the bright disc of sun called photosphere.

The Partial eclipse occurs when the sun, moon and earth are not exactly in line. Therefore the moon covers the small part of the sun.

The annular eclipse takes place when the moon's size is too small to completely cover the sun's photosphere which therefore appears as a bright ring around the sun.

LUNAR ECLIPSE:

The passing of earth directly between the moon and the sun results in a lunar eclipse. Like the solar eclipse the lunar eclipse also has three different types.
A total lunar eclipse occurs when the whole moon passes through the umbra. In case of partial eclipse the entire moon passes through the penumbra and only part of it passes through the umbra. And in a penumbral eclipse the moon passes through only the penumbra.

19. DAY AND NIGHT AND THEIR VARIATION

The rotation of the earth is responsible for the day and night varaitions.While rotating the half of the earth faces the sunlight while the other half faces away from the sun. The hemisphere of the earth that faces the sun has day time while the hemisphere that faces away from the sun has night time.
Earth completes its rotation in 24 hours and during this 24 hour duration the variation between day and night occurs.

20. ENERGY

Energy, capacity of matter to perform work as the result of its motion or its position in relation to forces acting on it. Energy associated with motion is known as kinetic energy, and energy related to position is called potential energy. Thus, a swinging pendulum has maximum potential energy at the terminal points; at all intermediate positions it has both kinetic and potential energy in varying proportions. Energy exists in various forms, including mechanical (see Mechanics), thermal (see Thermodynamics), chemical (see Chemical Reaction), electrical (see Electricity), radiant (see Radiation), and atomic (see Nuclear Energy). All forms of energy are interconvertible by appropriate processes. In the process of transformation either kinetic or potential energy may be lost or gained, but the sum total of the two remains always the same.

A weight suspended from a cord has potential energy due to its position, inasmuch as it can perform work in the process of falling. An electric battery has potential energy in chemical form. A piece of magnesium has potential energy stored in chemical form that is expended in the form of heat and light if the magnesium is ignited. If a gun is fired, the potential energy of the gunpowder is transformed into the kinetic energy of the moving projectile. The kinetic mechanical energy of the moving rotor of a dynamo is changed into kinetic electrical energy by electromagnetic induction. All forms of energy tend to be transformed into heat, which is the most transient form of energy. In mechanical devices energy not expended in useful work is dissipated in frictional heat, and losses in electrical circuits are largely heat losses.

Empirical observation in the 19th century led to the conclusion that although energy can be transformed, it cannot be created or destroyed. This concept, known as the conservation of energy, constitutes one of the basic principles of classical mechanics. The principle, along with the parallel principle of conservation of matter, holds true only for phenomena involving velocities that are small compared with the velocity of light. At higher velocities close to that of light, as in nuclear reactions, energy and matter are interconvertible (see Relativity). In modern physics the two concepts, the conservation of energy and of mass, are thus unified

21. SOURCES AND RESOURCES OF ENERGY

The ability to do work is energy. The significant sources of energy are coal, gas, wood and oil. These sources area so called the primary sources of energy. Electricity which is regarded as a secondary source of energy is produced by theses primary sources. For the production of electricity the sources of nuclear fission, sunlight and water are also employed as the primary sources.

CONVENTIONAL AND NON CONVENTIONAL SOURCES OF ENERGY:

The sources of energy that are used for the objective of power generation are called conventional sources of energy. Whereas the sources that are not utilized for the factor of electricity generation are included in the category of non conventional sources. However these categories change with time like once the nuclear energy was considered under a non conventional source of energy but with the modern discoveries now it is considered to be an important source of energy.

FORMS OF ENERGY:

COAL:
Coal is considered to be an important source of energy .Almost 30 % of world's power production is dependent on this form. It is a fossil fuel .The organic matter of plants is buried in rocks and soils. The pressure and heat changed this organic material to peat, lignite and then coal.

PETROLEUM:
Another essential form of energy is the use of petroleum. It is also a fossil fuel and a crude oil. Through the process of fractional distillation, the constituents of petroleum are separated .Major oil producing countries are USA, SAUDI ARABIA, ALGERIA and IRAN, KUWAIT.

NATURAL GAS:
Gas is a cheap source of energy and is an organic matter .It is also used for power generation and its major constituent elements are methane, ethane, propane and other hydrocarbons.
The natural gas is abundantly used in Pakistan especially.

HYDROELECTRIC POWER GENERATION:
Falling water in the mountainous areas is used as a source of mechanical energy to rotate turbines and generate electricity. The process of electromagnetic induction is used for this purpose.

SOLAR ENERGY:
The solar energy is used in photoelectric cells. When light strike certain heavy metals electricity is produced. Saudi Arabia, France and other European countries are utilizing this significant source of energy to heat buildings, power cars and other communication systems.

NUCLEAR ENERGY:
This form of energy is now being used by the countries like USA<UK<CANADA. In this source of energy is usually released in the form of heat which is first used to heat water and get steam. This steam is used to run a turbine which in return generates heat.

22. ENERGY CONSERVATION

By energy conservation it is meant that energy can neither be created nor destroyed but only converted into other forms. This principle infact is also known as the "law of conservation of energy".

For instance a pendulum that moves to and fro with the bob changing its speed from maximum to the lowest. In this process the kinetic energy is greatest at the lowest point of the swing and zero at the highet.howver the potential energy is maximum at the highest point and is zero at the lowest. This shows that kinetic energy changes to another form of energy .i.e. potential energy.

Similarly in every production of work all forms of energy generate in this manner. Heat energy is converted to light energy in some cases whereas in some the chemical energy is changed to the potential energy

23. CERAMICS

Ceramics include a vast variety of inorganic, non metallic materials which require high temperature heating for preparation. The most famous forms of ceramic are pottery, bricks, tiles and sanitary ware. Infact the ceramics have a wide variety of usage in most sectors especially in the industrial sector.

RAW MATERIAL:

The raw material available for ceramic preparation is usually the clay which is found beneath the top soil .This material is formed through the break down process of rocks affected from the chemical proceeds and the weather conditions. This raw material is very cheap and can be found in huge quantity.
This raw material when mixed with water can be shaped and molded in any form according to the requirements or usage. It becomes hard at heating thus making it more durable and strong.

KINDS:

Three kinds of pottery are in use at the present times.

1. EARTHENWARE
2. STONEWARE
3. PORCELAIN

EARTHENWARE: One of the cheapest and common styles of pottery is earthenware. It is easy to prepare and is heated at low temperature. However it is not durable and is dull, porous and absorbent.

STONEWARE: This kind of pottery is extremely hard and strong as is employed in the making of heavy dishes or jugs, sanitary wares etc.Howver it can hold liquids and is semi virtuous and glass like unlike earthenware.

PORCELAIN: This is the finest and the most refined form of pottery. It is translucent and strong light can shine through it.Porcealin is also called chinaware as the original technique originated from China.

24. PLASTICS

Plastics are man-made materials. Plastics have taken the place of traditional materials like woods and metals.
Plastics differ from other materials largely because of the size of their molecules. Most materials have molecules made up of less than 300 atoms, plastics contain thousands of atoms. We call them Macromolecules.

Some plastics are derived from natural substances such as animals, insects and plants but most are man-made. These are named Synthetic Plastics.

Most synthetic plastics come from crude oil but coal and natural gas is also used.
When crude oil is refined gasses are given off. The gasses are broken down into Monomers. These are chemical substances consisting of a single molecule. Thousands of these are linked together in a process called Polymerization to form new compounds called Polymers.

TYPES:

There are two main types of plastics and these are named Thermoplastics and Thermosetting Plastics.

Thermoplastics are made up of lines of molecules with few cross linkages. This allows them to soften when heated and to be bent into a variety of shapes and forms. They become stiff and solid again when cold. This process can be repeated many times

Thermosetting Plastics are made up of lines of molecules which are heavily cross linked. It creates a rigid molecular structure. They may be heated the first time and shaped but they become permanently stiff and solid. They cannot be reshaped again.

Plastic Memory Each time a plastic is reheated it will attempt to return to its original flat shape unless it has been over heated or damaged. This is called a plastic memory.

25. SEMI CONDUCTORS

Semi-conductors are materials with an electrical conductivity that increases with increasing temperature, a trend that is opposite to that of metals. Semi-conductors characteristically have a band gap between the valence and conduction bands that is smaller than that found in the insulators. The reason the conductivity increases is because as the temperature increases more electrons become thermally excited and are able to jump the band gap between the valence and conduction band. An example of this is silicon.

N-Type Conductivity:

When a foreign atom with an excess of electrons is added to a pure semi-conductor, the result is an n-type semi-conductor, so named because the charge carriers are negative. This increases the conductivity because a donor band, which is filled with electrons, is introduced near to the conduction band in the band gap. This greatly decreases the band gap which the electrons must jump. Therefore, more electrons are able to get to the conduction band and hence a greater conductivity is the result. An example of an n-type semi-conductor is germanium doped with phosphorous.

P-Type Conductivity:

When foreign atoms with less than 2N electrons are added, the result is a p-type semi-conductor, so called because the charge carrier is a positive hole. The foreign atoms create an acceptor band very close to the valence band that is empty. The result is that the band gap is deceased between a full and empty band. Electrons are then able to easily jumpfromthe valence band into the acceptor bands where they are trapped creating positive holes in the valence band. These positive create a means for the electrons to move within the valence band, thus increasing the conductivity.

26. RADIO

Radio is based on the principle that electrical signals have the capacity of travelling without the wire. Radio signals are carried by electromagnetic waves which travel through space at a speed of light.

The sound waves enter through a microphone in which a coil, a metal ribbon vibrates to change sound in an electric current. This signal of sound combines with a carrier signal which is at higher radio frequncy.The carrier is modulated by audio frequency signal. This modulated carrier signal get transfers to transmitting aerial where radio waves are emitted in all directions.

The received waves are fed into a radio frequency amplifier to stengthen; they pass to a detector which separates the audio frequency signal from the carrier wave. The currents that are obtained age identical to those that left the microphone at the broadcasting station. They are amplified and fed to the loudspeaker. The loudspeaker acts like a microphone, similar to the process that produces sounds like the original sound.

I -INTRODUCTION

Radio, system of communication employing electromagnetic waves propagated through space. Because of their varying characteristics, radio waves of different lengths are employed for different purposes and are usually identified by their frequency. The shortest waves have the highest frequency, or number of cycles per second; the longest waves have the lowest frequency, or fewest cycles per second. In honor of the German radio pioneer Heinrich Hertz, his name has been given to the cycle per second (hertz, Hz); 1 kilohertz (kHz) is 1000 cycles per sec, 1 megahertz (MHz) is 1 million cycles per sec, and 1 gigahertz (GHz) is 1 billion cycles per sec. Radio waves range from a few kilohertz to several gigahertz. Waves of visible light are much shorter. In a vacuum, all electromagnetic waves travel at a uniform speed of about 300,000 km (about 186,000 mi) per second. For electromagnetic waves other than radio.

Radio waves are used not only in radio broadcasting but also in wireless telegraphy, two-way communication for law enforcement, telephone transmission, wireless Internet, television, radar, navigational systems, GPS, and space communication. In the atmosphere, the physical characteristics of the air cause slight variations in velocity, which are sources of error in such radio-communications systems as radar. Also, storms or electrical disturbances produce anomalous phenomena in the propagation of radio waves.

Because electromagnetic waves in a uniform atmosphere travel in straight lines and because the earth's surface is approximately spherical, long-distance radio communication is made possible by the reflection of radio waves from the ionosphere. Radio waves shorter than about 10 m (about 33 ft) in wavelength—designated as very high, ultrahigh, and superhigh frequencies (VHF, UHF, and SHF)—are usually not reflected by the ionosphere; thus, in normal practice, such very short waves are received only within line-of-sight distances. Wavelengths shorter than a few centimeters are absorbed by water droplets or clouds; those shorter than 1.5 cm (0.6 in) may be absorbed selectively by the water vapor present in a clear atmosphere.

A typical radio communication system has two main components, a transmitter and a receiver. The transmitter generates electrical oscillations at a radio frequency called the carrier frequency. Either the amplitude or the frequency itself may be modulated to vary the carrier wave. An amplitude-modulated signal consists of the carrier frequency plus two sidebands resulting from the modulation. Frequency modulation produces more than one pair of sidebands for each modulation frequency. These produce the complex variations that emerge as speech or other sound in radio broadcasting, and in the alterations of light and darkness in television broadcasting.

II -TRANSMITTER

Essential components of a radio transmitter include an oscillation generator for converting commercial electric power into oscillations of a predetermined radio frequency; amplifiers for increasing the intensity of these oscillations while retaining the desired frequency; and a transducer for converting the information to be transmitted into a varying electrical voltage proportional to each successive instantaneous intensity. For sound transmission a microphone is the transducer; for picture transmission the transducer is a photoelectric device.

Other important components of the radio transmitter are the modulator, which uses these proportionate voltages to control the variations in the oscillation intensity or the instantaneous frequency of the carrier, and the antenna, which radiates a similarly modulated carrier wave. Every antenna has some directional properties, that is, it radiates more energy in some directions than in others, but the antenna can be modified so that the radiation pattern varies from a comparatively narrow beam to a comparatively even distribution in all directions; the latter type of radiation is employed in broadcasting.

The particular method of designing and arranging the various components depends on the effects desired. The principal criteria of a radio in a commercial or military airplane, for example, are light weight and intelligibility; cost is a secondary consideration, and fidelity of reproduction is entirely unimportant. In a commercial broadcasting station, on the other hand, size and weight are of comparatively little importance; cost is of some importance; and fidelity is of the utmost importance, particularly for FM stations; rigid control of frequency is an absolute necessity. In the U.S., for example, a typical commercial station broadcasting on 1000 kHz is assigned a bandwidth of 10 kHz by the Federal Communications Commission, but this width may be used only for modulation; the carrier frequency itself must be kept precisely at 1000 kHz, for a deviation of one-hundredth of 1 percent would cause serious interference with even distant stations on the same frequency.

A -Oscillators

In a typical commercial broadcasting station the carrier frequency is generated by a carefully controlled quartz-crystal oscillator. The fundamental method of controlling frequencies in most radio work is by means of tank circuits, or tuned circuits, that have specific values of inductance and capacitance, and that therefore favor the production of alternating currents of a particular frequency and discourage the flow of currents of other frequencies. In cases where the frequency must be extremely stable, however, a quartz crystal with a definite natural frequency of electrical oscillation is used to stabilize the oscillations. The oscillations are actually generated at low power by an electron tube and are amplified in a series of power amplifiers that act as buffers to prevent interaction of the oscillator with the other components of the transmitter, because such interaction would alter the frequency.

The crystal is shaped accurately to the dimensions required to give the desired frequency, which may then be modified slightly by adding a condenser to the circuit to give the exact frequency desired. In a well-designed circuit, such an oscillator does not vary by more than one-hundredth of 1 percent in frequency. Mounting the crystal in a vacuum at constant temperature and stabilizing the supply voltages may produce a frequency stability approaching one-millionth of 1 percent. Crystal oscillators are most useful in the ranges termed very low frequency, low frequency, and medium frequency (VLF, LF, and MF). When frequencies higher than about 10 MHz must be generated, the master oscillator is designed to generate a medium frequency, which is then doubled as often as necessary in special electronic circuits. In cases where rigid frequency control is not required, tuned circuits may be used with conventional electron tubes to generate oscillations up to about 1000 MHz, and reflex klystrons are used to generate the higher frequencies up to 30,000 MHz. Magnetrons are substituted for klystrons when even larger amounts of power must be generated.

B -Modulation

Modulation of the carrier wave so that it may carry impulses is performed either at low level or high level. In the former case the audio-frequency signal from the microphone, with little or no amplification, is used to modulate the output of the oscillator, and the modulated carrier frequency is then amplified before it is passed to the antenna; in the latter case the radio-frequency oscillations and the audio-frequency signal are independently amplified, and modulation takes place immediately before the oscillations are passed to the antenna. The signal may be impressed on the carrier either by frequency modulation (FM) or amplitude modulation (AM).

The simplest form of modulation is keying, interrupting the carrier wave at intervals with a key or switch used to form the dots and dashes in continuous-wave radiotelegraphy.

The carrier wave may also be modulated by varying the amplitude, or strength, of the wave in accordance with the variations of frequency and intensity of a sound signal, such as a musical note. This form of modulation, AM, is used in many radiotelephony services including standard radiobroadcasts. AM is also employed for carrier current telephony, in which the modulated carrier is transmitted by wire, and in the transmission of still pictures by wire or radio.

In FM the frequency of the carrier wave is varied within a fixed range at a rate corresponding to the frequency of a sound signal. This form of modulation, perfected in the 1930s, has the advantage of yielding signals relatively free from noise and interference arising from such sources as automobile-ignition systems and thunderstorms, which seriously affect AM signals. As a result, FM broadcasting is done on high-frequency bands (88 to 108 MHz), which are suitable for broad signals but have a limited reception range.

Carrier waves can also be modulated by varying the phase of the carrier in accordance with the amplitude of the signal. Phase modulation, however, has generally been limited to special equipment. The development of the technique of transmitting continuous waves in short bursts or pulses of extremely high power introduced the possibility of yet another form of modulation, pulse-time modulation, in which the spacing of the pulses is varied in accordance with the signal.

The information carried by a modulated wave is restored to its original form by a reverse process called demodulation or detection. Radio waves broadcast at low and medium frequencies are amplitude modulated. At higher frequencies both AM and FM are in use; in present-day commercial television, for example, the sound may be carried by FM, while the picture is carried by AM. In the super high-frequency range (above the ultrahigh-frequency range), in which broader bandwidths are available, the picture also may be carried by FM.

Digital radio (also called HD or high-definition radio) processes sounds into patterns of numbers instead of into patterns of electrical waves and can be used for both FM and AM broadcasts. The sound received by a radio listener is much clearer and virtually free from interference. The signals can be used to provide additional services, multiple channels, and interactive features. Satellite radio is also a form of digital radio but the signal is broadcast from communication satellites in orbit around Earth and not from local broadcast towers.

C -Antennas

The antenna of a transmitter need not be close to the transmitter itself. Commercial broadcasting at medium frequencies generally requires a very large antenna, which is best located at an isolated point far from cities, whereas the broadcasting studio is usually in the heart of the city. FM, television, and other very-high-frequency broadcasts must have very high antennas if appreciably long range is to be achieved, and it may not be convenient to locate such a high antenna near the broadcasting studio. In all such cases, the signals may be transmitted by wires. Ordinary telephone lines are satisfactory for most commercial radio broadcasts; if high fidelity or very high frequencies are required, coaxial or fiber optic cables are used.

III -RECEIVERS

The essential components of a radio receiver are an antenna for receiving the electromagnetic waves and converting them into electrical oscillations; amplifiers for increasing the intensity of these oscillations; detection equipment for demodulating; a speaker for converting the impulses into sound waves audible by the human ear (and in television a picture tube for converting the signal into visible light waves); and, in most radio receivers, oscillators to generate radio-frequency waves that can be ―mixed‖ with the incoming waves.

The incoming signal from the antenna, consisting of a radio-frequency carrier oscillation modulated by an audio frequency or video-frequency signal containing the impulses, is generally very weak. The sensitivity of some modern radio receivers is so great that if the antenna signal can produce an alternating current involving the motion of only a few hundred electrons, this signal can be detected and amplified to produce an intelligible sound from the speaker. Most radio receivers can operate quite well with an input from the antenna of a few millionths of a volt. The dominant consideration in receiver design, however, is that very weak desired signals cannot be made useful by amplifying indiscriminately both the desired signal and undesired radio noise (see Noise below). Thus, the main task of the designer is to assure preferential reception of the desired signal.

Most modern radio receivers are of the super heterodyne type in which an oscillator generates a radio-frequency wave that is mixed with the incoming wave, thereby producing a radio-frequency wave of lower frequency; the latter is called intermediate frequency. To tune the receiver to different frequencies, the frequency of the oscillations is changed, but the intermediate frequency always remains the same (at 455 kHz for most AM receivers and at 10.7 MHz for most FM receivers). The oscillator is tuned by altering the capacity of the capacitor in its tank circuit; the antenna circuit is similarly tuned by a capacitor in its circuit. One or more stages of intermediate-frequency amplification are included in all receivers; in addition, one or more stages of radiofrequency amplification may be included.

Auxiliary circuits such as automatic volume control (which operates by rectifying part of the output of one amplification circuit and feeding it back to the control element of the same circuit or of an earlier one) are usually included in the intermediate-frequency stage. The detector, often called the second detector, the mixer being called the first detector, is usually simply a diode acting as a rectifier, and produces an audio-frequency signal. FM waves are demodulated or detected by circuits known as discriminators or radio-detectors that translate the varying frequencies into varying signal amplitudes.

Digital and satellite radio require special receivers that can change a digital signal into analog sound. The digital signal can carry additional information that can be displayed on a screen on the radio. The title of a music track and the artist can be provided, for example. Some radios can even record songs in MP3 format.

A -Amplifiers

Radio-frequency and intermediate-frequency amplifiers are voltage amplifiers, increasing the voltage of the signal. Radio receivers may also have one or more stages of audio-frequency voltage amplification. In addition, the last stage before the speaker must be a stage of power amplification. A high-fidelity receiver contains both the tuner and amplifier circuits of a radio. Alternatively, a high-fidelity radio may consist of a separate audio amplifier and a separate radio tuner.

The principal characteristics of a good radio receiver are high sensitivity, selectivity, fidelity, and low noise. Sensitivity is primarily achieved by having numerous stages of amplification and high amplification factors, but high amplification is useless unless reasonable fidelity and low noise can be obtained. The most sensitive receivers have one stage of tuned radio-frequency amplification. Selectivity is the ability of the receiver to obtain signals from one station and reject signals from another station operating on a nearby frequency. Excessive selectivity is not desirable, because a bandwidth of many kilohertz is necessary in order to receive the high-frequency components of the audio-frequency signals. A good broadcast-band receiver tuned to one station has a zero response to a station 20 kHz away. The selectivity depends principally on the circuits in the intermediate-frequency stage.

B -High-Fidelity Systems
Fidelity is the equality of response of the receiver to various audio-frequency signals modulated on the carrier. Extremely high fidelity, which means a flat frequency response (equal amplification of all audio frequencies) over the entire audible range from about 20 Hz to 20 kHz, is extremely difficult to obtain. A high-fidelity system is no stronger than its weakest link, and the links include not only all the circuits in the receiver, but also the speaker, the acoustic properties of the room in which the speaker is located, and the transmitter to which the receiver is tuned. Most AM radio stations do not reproduce faithfully sounds below 100 Hz or above 5 kHz; FM stations generally have a frequency range of 50 Hz to 15 kHz, the upper limit being set by Federal Communications Commission regulations. Digital and satellite radio can provide even better high fidelity over a larger range of frequencies. Digital FM approaches the sound quality of CDs. Digital AM radio should be comparable to regular FM in sound quality.

C -Distortion
A form of amplitude distortion is often introduced to a radio transmission by increasing the relative intensity of the higher audio frequencies. At the receiver, a corresponding amount of high-frequency attenuation is applied. The net effect of these two forms of distortion is a net reduction in high-frequency background noise or static at the receiver. Many receivers are also equipped with user-adjustable tone controls so that the amplification of high and low frequencies may be adjusted to suit the listener's taste. Another source of distortion is cross modulation, the transfer of signals from one circuit to another through improper shielding. Harmonic distortion caused by nonlinear transfer of signals through amplification stages can often be significantly reduced by the use of negative-feedback circuitry that tends to cancel most of the distortion generated in such amplification stages.

D -Noise
Noise is a serious problem in all radio receivers. Several different types of noise, each characterized by a particular type of sound and by a particular cause, have been given names. Among these are hum, a steady low-frequency note (about two octaves below middle C) commonly produced by the frequency of the alternating-current power supply (usually 60 Hz) becoming impressed onto the signal because of improper filtering or shielding; hiss, a steady high-frequency note; and whistle, a pure high-frequency note produced by unintentional audio-frequency oscillation, or by beats. These noises can be eliminated by proper design and construction. Certain types of noise, however, cannot be eliminated. The most important of these in ordinary AM low-frequency and medium-frequency sets is static, caused by electrical disturbances in the atmosphere.

Static may be due to the operation of nearby electrical equipment (such as automobile and airplane engines), but is most often caused by lightning. Radio waves produced by such atmospheric disturbances can travel thousands of kilometers with comparatively little attenuation, and inasmuch as a thunderstorm is almost always occurring somewhere within a few thousand kilometers of any radio receiver, static is almost always present. Static affects FM receivers to a much smaller degree, because the amplitude of the intermediate waves is limited in special circuits before discrimination, and this limiting removes effects of static, which influences the signal only by superimposing a random amplitude modulation on the wave. Digital and satellite radio greatly reduces static.

Another basic source of noise is thermal agitation of electrons. In any conductor at a temperature higher than absolute zero, electrons are moving about in a random manner. Because any motion of electrons constitutes an electric current, this thermal motion gives rise to noise when amplification is carried too far. Such noise can be avoided if the signal received from the antenna is considerably stronger than the current caused by thermal agitation; in any case, such noise can be minimized by suitable design. A theoretically perfect receiver at ordinary temperatures can receive speech intelligibly when the signal power in the antenna is only 4 × 10-18 W (40 attowatts); in ordinary radio receivers, however, considerably greater signal strength is required.

E -Power Supply
A radio has no moving parts except the speaker cone, which vibrates within a range of a few thousandths of a centimeter, and so the only power required to operate the radio is electrical power to force electrons through the various circuits. When radios first came into general use in the 1920s, batteries operated most. Although batteries are used widely in portable sets today, a power supply from a power line has advantages, because it permits the designer more freedom in selecting circuit components. If the alternating-current (AC) power supply is 120 V, this current can be led directly to the primary coil of a transformer, and power with the desired voltage can be drawn off as desired from the secondary coils. This secondary current must be rectified and filtered before it can be used because transistors require direct current (DC) for proper operation. Electron tubes require DC for plate current; filaments may be heated either by DC or AC, but in the latter case hum may be created.

Transistorized radios do not require as high an operating DC voltage as did tube radios of the past, but power supplies are still needed to convert the AC voltage distributed by utility companies to DC, and to step up or step down the voltage to the required value, using transformers. Airplane and automobile radio sets that operate on 12 to 24 volts DC often contain circuits that convert the available DC voltage to AC, after which the voltage is stepped up or down to the required voltage level and again converted to DC by a rectifier . Airplane and automobile radio sets that operate on 6 to 24 volts DC always contain some such device for raising the voltage. The advent of transistors, integrated circuits, and other solid-state electronic devices, which are much smaller in size and require very little power, has today greatly reduced the use of vacuum tubes in radio, television, and other types of communications equipment and devices.

IV -HISTORY
Although many discoveries in the field of electricity were necessary to the development of radio, the history of radio really began in 1873, with the publication by the British physicist James Clerk Maxwell of his theory of electromagnetic waves.

A -Late 19th Century
Maxwell's theory applied primarily to light waves. About 15 years later the German physicist Heinrich Hertz actually generated such waves electrically. He supplied an electric charge to a capacitor, and then short-circuited the capacitor through a spark gap. In the resulting electric discharge the current surged past the neutral point, building up an opposite charge on the capacitor, and

then continued to surge back and forth, creating an oscillating electric discharge in the form of a spark. Some of the energy of this oscillation was radiated from the spark gap in the form of electromagnetic waves. Hertz measured several of the properties of these so-called Hertzian waves, including their wavelength and velocity.

The concept of using electromagnetic waves for the transmission of messages from one point to another was not new; the heliograph, for example, successfully transmitted messages via a beam of light rays, which could be modulated by means of a shutter to carry signals in the form of the dots and dashes of the Morse code. Radio has many advantages over light for this purpose, but these advantages were not immediately apparent. Radio waves, for example, can travel enormous distances; but microwaves (which Hertz used) cannot. Radio waves can be enormously attenuated and still be received, amplified, and detected; but good amplifiers were not available until the development of electron tubes. Although considerable progress was made in radiotelegraphy (for example, transatlantic communication was established in 1901), radiotelephony might never have become practical without the development of electronics. Historically, developments in radio and in electronics have been interdependent.

To detect the presence of electromagnetic radiation, Hertz used a loop of wire somewhat similar to a wire antenna. At about the same time the Anglo-American inventor David Edward Hughes discovered that a loose contact between a steel point and a carbon block would not conduct current, but that if electromagnetic waves were passed through the junction point, it conducted well. In 1879 Hughes demonstrated the reception of radio signals from a spark transmitter located some hundreds of meters away. In these experiments he conducted a current from a voltaic cell through a glass tube filled loosely with zinc and silver filings, which cohered when radio waves impinged on it. The British physicist Sir Oliver Joseph Lodge, in a device called the coherer, to detect the presence of radio waves, used the principle.

The coherer, after becoming conductive, could again be made resistant by tapping it, causing the metal particles to separate. Although far more sensitive than a wire loop in the absence of an amplifier, the coherer gave only a single response to sufficiently strong radio waves of varying intensities, and could thus be used for telegraphy but not for telephony.
The Italian electrical engineer and inventor Guglielmo Marconi is generally credited with being the inventor of radio. Starting in 1895 he developed an improved coherer and connected it to a rudimentary form of antenna, with its lower end grounded. He also developed improved spark oscillators, connected to crude antennas. The transmitter was modulated with an ordinary telegraph key.

The coherer at the receiver actuated a telegraphic instrument through a relay, which functioned as a crude amplifier. In 1896 he transmitted signals for a distance exceeding 1.6 km (more than 1 mi), and applied for his first British patent. In 1897 he transmitted signals from shore to a ship at sea 29 km (18 mi) away. In 1899 he established commercial communication between England and France that operated in all types of weather; early in 1901 he sent signals 322 km (200 mi), and later in the same year succeeded in sending a single letter across the Atlantic Ocean. In 1902 messages were regularly sent across the Atlantic, and by 1905 many ships were using radio for communications with shore stations. For his pioneer work in the field of wireless telegraphy, Marconi shared the 1909 Nobel Prize in physics with the German physicist Karl Ferdinand Braun.

During this time various technical improvements were being made. Tank circuits, containing inductance and capacitance, were used for tuning. Antennas were improved, and their directional properties were discovered and used. Transformers were used to increase the voltage sent to the antenna. Other detectors were developed to supplement the coherer with its clumsy tapper; among these were a magnetic detector that depended on the ability of radio waves to demagnetize steel wires; a bolometer that measured the rise in temperature of a fine wire when radio waves are passed through the wire; and the so-called Fleming valve, the forerunner of the thermionic tube, or vacuum tube.

B -20th Century
The modern vacuum tube traces its development to the discovery made by the American inventor Thomas Alva Edison that a current will flow between the hot filament of an incandescent lamp and another electrode placed in the same lamp, and that this current will flow in only one direction. The Fleming valve was not essentially different from Edison's tube. It was developed by the British physicist and electrical engineer Sir John Ambrose Fleming in 1904 and was the first of the diodes, or two-element tubes, used in radios. This tube was then used as a detector, rectifier, and limiter. A revolutionary advance, which made possible the science of electronics, occurred in 1906 when the American inventor Lee De Forest mounted a third element, the grid, between the filament and cathode of a vacuum tube. De Forest's tube, which he called an audio but which is now called a triode (three-element tube), was first used only as a detector, but its potentialities as an amplifier and oscillator were soon developed, and by 1915 wireless telephony had developed to such a point that communication was established between Virginia and Hawaii and between Virginia and Paris.

The rectifying properties of crystals were discovered in 1912 by the American electrical engineer and inventor Greenleaf Whittier Pickard, who pointed out that crystals can be used as detectors. This discovery gave rise to the so-called crystal sets popular about 1920. In 1912 the American electrical engineer Edwin Howard Armstrong discovered the regenerative circuit, by which part of the output of a tube is fed back to the same tube. This and certain other discoveries by Armstrong form the basis of many circuits in modern radio sets.

In 1902 the American electrical engineer Arthur Edwin Kennelly and the British physicist and electrician Oliver Heaviside, independently and almost simultaneously, announced the probable existence of a layer of ionized gas high in the atmosphere that affects the propagation of radio waves. This layer, formerly called the Heaviside or Kennelly-Heaviside layer, is one of several layers in the ionosphere. Although the ionosphere is transparent to the shortest radio wavelengths, it bends or reflects the longer waves. Because of this reflection, radio waves can be propagated far beyond the horizon. Propagation of radio waves in the ionosphere is strongly affected by time of day, season, and sunspot activity. Slight variations in the nature and altitude of the ionosphere, which can occur rapidly, can affect the quality of long-distance reception.

The ionosphere is also responsible for skip, the reception at a considerable distance of a signal that cannot be received at a closer point. This phenomenon occurs when the intervening ground has absorbed the ground ray and the ion spherically propagated ray is not reflected at an angle sufficiently steep to be received at short distances from the antenna.

C -Short-wave Radio

Although parts of the various radio bands—short-wave, long-wave, medium-wave, very-high frequency, and ultrahigh frequency— are allocated for a variety of purposes, the term short-wave radio generally refers to radiobroadcasts in the high-frequency range (3 to 30 MHz) beamed for long distances, especially in international communication. Microwave communication via satellite, however, provides signals with superior reliability and freedom from error.

Amateur, or ―ham,‖ radio is also commonly thought of as short-wave, although amateur operators have been allotted frequencies in the medium-wave band, the very-high-frequency band, and the ultrahigh-frequency band as well as the short-wave band. Certain of these frequencies have restrictions designed to make them available to maximum numbers of users. During the rapid development of radio after World War I, amateur operators executed such spectacular feats as the first transatlantic radio contact (1921). They have also provided valuable voluntary assistance during emergencies when normal communications are disrupted. Amateur radio organizations have launched a number of satellites piggyback with regular launches by the United States, the former Soviet Union, and the European Space Agency.

These satellites are usually called Oscar, for Orbiting Satellites Carrying Amateur Radio. The first, Oscar 1, orbited in 1961, was also the first nongovernmental satellite; the fourth, in 1965, provided the first direct-satellite communications between the U.S. and the Soviet Union. More than 1.5 million people worldwide were licensed amateur radio operators in the early 1980s.

The ability to webcast radio programs over the Internet had a major impact on shortwave broadcasting. In the early 2000s the BBC dropped their shortwave radio service to the United States, Canada, Australia, and other developed countries since their programs were available through computers over the World Wide Web. The widespread use of personal computers with Internet access to chat groups and personal Web pages also replaced some of the hobby aspects of amateur radio in popularity.

D -Radio today

Immense developments in radio communication technology after World War II helped make possible space exploration, most dramatically in the Apollo moon-landing missions (1969-72). Sophisticated transmitting and receiving equipment was part of the compact, very-high-frequency, communication system on board the command modules and the lunar modules. The system performed voice and ranging functions simultaneously, calculating the distance between the two vehicles by measuring the time lapse between the transmission of tones and the reception of the returns. The voice signals of the astronauts were also transmitted simultaneously around the world by a communications network.

In the 1990s cellular radio telephones (cell phones) became one of the most important and widespread uses of radio communication. By the early 21st century, billions of people worldwide had access to telephone service with lightweight portable cell phones capable of communicating worldwide through radio relays and satellite links. Cell phones have become particularly important in developing countries where landlines for telephones often do not exist outside of large cities. In remote rural areas an individual who owns a cell phone may charge a small fee to let others use the phone service. Such phone service can have a major economic impact in impoverished regions, permitting access to banking services, providing information on prices of crops, and creating small-business contacts.

Digital and satellite radio also greatly expanded the possibilities of radio. Not only does digital radio provide superior sound quality, but it permits such additional services as multiple audio-programming channels, on-demand audio services, and interactive features, as well as targeted advertising. Wireless Internet allows users of computers and portable media devices to access the World Wide Web from all kinds of locations. Personal digital assistants (PDAs) also use radio to access e-mail and other services, including GPS information from satellites. The transition to digital television is expected to free up a large part of the radio spectrum previously used to broadcast analog television. These frequencies may be available for many more wireless uses in the future

27. TELEVISION

The TV works through the electromagnetic signals that are transmitted, received and converted back to original patterns. The sound transmission is very much similar to radio. In picture transmission the fundamental component is the camera which is responsible for changing the image into electrical impulses...The cathode ray tube at the other end converts the pattern of electrical impulses into visible images.

Inside the TV camera an illuminated plate emits electrons. The electrons travel to a plate which is called a target plate. The electrical pattern that produces afterwards is transmitted to the transmitter where the synchronizing pulses are added. Before the final output is fed to the transmitting aerials the sound signal is added.

For transmission VHF and UHF frequencies are used .The receiver is based on the super heterodyne principle, the sound, the vision are received at the separate intermediate frequency amplifiers, detectors and output stages. The electron beam is made to scan the screen of the cathode, ray tube and in step with the beam in the TV comers' picture is then received on the screen

28. TELEPHONES

The telephone consists of coils of fine insulated wire that is wound around a permanent horse shoe magnet. A soft iron disc diaphragm is held near the end of the magnet. The magnet lines of force gather in this disc. When the disc is thrown into vibration by a human voice, the number of lines of force passing through the coil changes and a fluctuating current is induced. At the receiving end the terminals over the coil wound over the ogles of another horse shoe magnet produces the similar vibrations that are produced at the transmitting end and thus helps in producing the sound.

29. CAMERA

Equipment for taking photographs which usually consists of a lightproof box with a lens at one end and light-sensitive film at the other.

Photography is undoubtedly one of the most important inventions in history -- it has truly transformed how people conceive of the world. Now we can "see" all sorts of things that are actually many miles -- and years -- away from us. Photography lets us capture moments in time and preserve them for years to come.

The basic technology that makes all of this possible is fairly simple. A still film camera is made of three basic elements: an optical element (the lens), a chemical element (the film) and a mechanical element (the camera body itself). As we'll see, the only trick to photography is calibrating and combining these elements in such a way that they record a crisp, recognizable image.

EVERY camera has these basic parts.
This first and main part is called the body.

The second part is the shutter which might be located in the lens (leaf shutter) or it might be located right in front of the film (focal plane shutter).

The shutter controls WHEN the light enters the camera and for how long it enters. The shutter in the lens is often faster and quieter, but makes changing the lens difficult. The shutter in front of the film allows for easy lens removal, but is often loud and slow. A good camera will have some way of adjusting the time the shutter is open plus there has to be some type of release for the shutter.

The lens lets in light. The larger the lens the more light. The lens also effects how large the image appears based on the focal length of the lens. The aperture is located in the lens and is a set of leaf like piece of metal that can change the size of the hole that lets in light. We consider the lens to be part of the shutter as we do not actually need a lens to focus an image if we have a small enough hole to let in the light.

Finally, the third part is film holder inside the camera. This must have some attachment that allows for the film to be moved which can either be a lever or a motor.

30. LASERS

Laser light has several features that are significantly different from white light. To begin with, light from most sources spreads out as it travels, so that much less light hits a given area as the distance from the light source increases. Laser light travels as a parallel beam and spreads very little.

Furthermore, laser light is monochromatic and coherent. White light is a jumble of colored light waves. Each color has a different wavelength. If all the wavelengths but one are filtered out, the remaining light is monochromatic. If these waves are all parallel to one another, they are also coherent: the waves travel in a definite phase relationship with one another. In the case of laser light, the wave crests coincide and the troughs coincide. The waves all reinforce one another. It is the monochromaticity and coherency of laser light that makes it ideal for recording data on optical media such as a CD as well as use as a light source for long haul fiber-optic communications.

The laser uses a process called stimulated emission to amplify light waves. (One method of amplification of an electromagnetic beam is to produce additional waves that travel in step with that beam.) A substance normally gives off light by spontaneous emission. One of the electrons of an atom absorbs energy. While it possesses this energy, the atom is in an excited state. If the electron gives off this excess energy (in the form of electromagnetic radiation such as light) with no outside impetus, spontaneous emission has occurred.

If a wave emitted by one excited atom strikes another, it stimulates the second atom to emit energy in the form of a second wave that travels parallel to and in step with the first wave. This stimulated emission results in amplification of the first wave. If the two waves strike other excited atoms, a large coherent beam builds up. But if they strike unexcited atoms, they are simply absorbed, and the amplification is then lost. In the case of normal matter on Earth, the great majority of atoms are not excited. As more than the usual number of atoms become excited, the probability increases that stimulated emission rather than absorption will take place.

Physicist Gordon Gould invented the laser in 1958. The first working model was built in 1960 by T.H. Maiman. It contained a synthetic, cylindrical ruby with a completely reflecting silver layer on one end and a partially reflecting silver layer on the other. Ruby is composed of aluminum oxide with chromium impurities. The chromium atoms absorb blue light and become excited; they then drop first to a metastable level and finally to the ground (unexcited) state, giving off red light. Light from a flash lamp enters the ruby and excites most of the chromium atoms, many of which fall quickly to the metastable level. Some atoms then emit red light and return to the ground state. The light waves strike other excited chromium atoms, stimulating them to emit more red light. The beam bounces back and forth between the silvered ends until it gains enough energy to burst through the partially silvered end as laser light. When most of the chromium atoms are back in the ground state, they absorb light, and the lasing action stops. In continuous-wave lasers, such as the helium-neon laser, electrons emit light by jumping to a lower excited state, forming a new atomic population that does not absorb laser light, rather than to the ground state.

I -INTRODUCTION

Laser, a device that produces and amplifies light. The word laser is an acronym for Light Amplification by Stimulated Emission of Radiation. Laser light is very pure in color, can be extremely intense, and can be directed with great accuracy. Lasers are used in many modern technological devices including bar code readers, compact disc (CD) players, and laser printers. Lasers can generate light beyond the range visible to the human eye, from the infrared through the X-ray range. Masers are similar devices that produce and amplify microwaves.

II -PRINCIPLES OF OPERATION

Lasers generate light by storing energy in particles called electrons inside atoms and then inducing the electrons to emit the absorbed energy as light. Atoms are the building blocks of all matter on Earth and are a thousand times smaller than viruses. Electrons are the underlying source of almost all light.
Light is composed of tiny packets of energy called photons. Lasers produce coherent light: light that is monochromatic (one color) and whose photons are ―in step‖ with one another.

A -Excited Atoms

At the heart of an atom is a tightly bound cluster of particles called the nucleus. This cluster is made up of two types of particles: protons, which have a positive charge, and neutrons, which have no charge. The nucleus makes up more than 99.9 percent of the atom‵s mass but occupies only a tiny part of the atom‵s space. Enlarge an atom up to the size of Yankee Stadium and the equally magnified nucleus is only the size of a baseball.

Electrons, tiny particles that have a negative charge, whirl through the rest of the space inside atoms. Electrons travel in complex orbits and exist only in certain specific energy states or levels . Electrons can move from a low to a high energy level by absorbing energy. An atom with at least one electron that occupies a higher energy level than it normally would is said to be excited. An atom can become excited by absorbing a photon whose energy equals the difference between the two energy levels. A photon‵s energy, color, frequency, and wavelength are directly related: All photons of a given energy are the same color and have the same frequency and wavelength.
Usually, electrons quickly jump back to the low energy level, giving off the extra energy as light (see Photoelectric Effect). Neon signs and fluorescent lamps glow with this kind of light as many electrons independently emit photons of different colors in all directions.

B -Stimulated Emission

Lasers are different from more familiar sources of light. Excited atoms in lasers collectively emit photons of a single color, all traveling in the same direction and all in step with one another. When two photons are in step, the peaks and troughs of their waves line up. The electrons in the atoms of a laser are first pumped, or energized, to an excited state by an energy source. An excited atom can then be ―stimulated‖ by a photon of exactly the same color (or, equivalently, the same wavelength) as the photon this atom is about to emit spontaneously. If the photon approaches closely enough, the photon can stimulate the excited atom to immediately emit light that has the same wavelength and is in step with the photon that interacted with it. This stimulated emission is the key to laser operation. The new light adds to the existing light, and the two photons go on to stimulate other excited atoms to give up their extra energy, again in step. The phenomenon snowballs into an amplified, coherent beam of light: laser light.

In a gas laser, for example, the photons usually zip back and forth in a gas-filled tube with highly reflective mirrors facing inward at each end. As the photons bounce between the two parallel mirrors, they trigger further stimulated emissions and the light gets brighter and brighter with each pass through the excited atoms. One of the mirrors is only partially silvered, allowing a small amount of light to pass through rather than reflecting it all. The intense, directional, and single-colored laser light finally escapes through this slightly transparent mirror. The escaped light forms the laser beam.

Albert Einstein first proposed stimulated emission, the underlying process for laser action, in 1917. Translating the idea of stimulated emission into a working model, however, required more than four decades. The working principles of lasers were outlined by the American physicists Charles Hard Townes and Arthur Leonard Schawlow in a 1958 patent application. (Both men won Nobel Prizes in physics for their work, Townes in 1964 and Schawlow in 1981). The patent for the laser was granted to Townes and Schawlow, but it was later challenged by the American physicist and engineer Gordon Gould, who had written down some ideas and coined the word laser in 1957. Gould eventually won a partial patent covering several types of laser. In 1960 American physicist Theodore Maiman of Hughes Aircraft Corporation constructed the first working laser from a ruby rod.

III -TYPES OF LASERS

Lasers are generally classified according to the material, called the medium, they use to produce the laser light. Solid-state, gas, liquid, semiconductor, and free electron are all common types of lasers.

A -Solid-State Lasers

Solid-state lasers produce light by means of a solid medium. The most common solid laser media are rods of ruby crystals and neodymium-doped glasses and crystals. The ends of the rods are fashioned into two parallel surfaces coated with a highly reflecting nonmetallic film. Solid-state lasers offer the highest power output. They are usually pulsed to generate a very brief burst of light. Bursts as short as 12 × 10-15 sec have been achieved. These short bursts are useful for studying physical phenomena of very brief duration. One method of exciting the atoms in lasers is to illuminate the solid laser material with higher-energy light than the laser produces. This procedure, called pumping, is achieved with brilliant strobe light from xenon flash tubes, arc lamps, or metal-vapor lamps.

B -Gas Lasers

The lasing medium of a gas laser can be a pure gas, a mixture of gases, or even metal vapor. The medium is usually contained in a cylindrical glass or quartz tube. Two mirrors are located outside the ends of the tube to form the laser cavity. Gas lasers can be pumped by ultraviolet light, electron beams, electric current, or chemical reactions. The helium-neon laser is known for its color purity and minimal beam spread. Carbon dioxide lasers are very efficient at turning the energy used to excite their atoms into laser light. Consequently, they are the most powerful continuous wave (CW) lasers—that is, lasers that emit light continuously rather than in pulses.

C -Liquid Lasers

The most common liquid laser media are inorganic dyes contained in glass vessels. They are pumped by intense flash lamps in a pulse mode or by a separate gas laser in the continuous wave mode. Some dye lasers are tunable, meaning that the color of the laser light they emit can be adjusted with the help of a prism located inside the laser cavity.

D -Semiconductor Lasers

Semiconductor lasers are the most compact lasers. Gallium arsenide is the most common semiconductor used. A typical semiconductor laser consists of a junction between two flat layers of gallium arsenide. One layer is treated with an impurity whose atoms provide an extra electron, and the other with an impurity whose atoms are one electron short. Semiconductor lasers are pumped by the direct application of electric current across the junction. They can be operated in the continuous wave mode with better than 50 percent efficiency. Only a small percentage of the energy used to excite most other lasers is converted into light.

Scientists have developed extremely tiny semiconductor lasers, called quantum-dot vertical-cavity surface-emitting lasers. These lasers are so tiny that more than a million of them can fit on a chip the size of a fingernail.
Common uses for semiconductor lasers include compact disc (CD) players and laser printers. Semiconductor lasers also form the heart of fiber-optics communication systems.

E -Free Electron Lasers

Free electron lasers employ an array of magnets to excite free electrons (electrons not bound to atoms). First developed in 1977, they are now becoming important research instruments. Free electron lasers are tunable over a broader range of energies than dye lasers. The devices become more difficult to operate at higher energies but generally work successfully from infrared through ultraviolet wavelengths. Theoretically, electron lasers can function even in the X-ray range.

The free electron laser facility at the University of California at Santa Barbara uses intense far-infrared light to investigate mutations in DNA molecules and to study the properties of semiconductor materials. Free electron lasers should also eventually become capable of producing very high-power radiation that is currently too expensive to produce. At high power, near-infrared beams from a free electron laser could defend against a missile attack.

IV -LASER APPLICATIONS

The use of lasers is restricted only by imagination. Lasers have become valuable tools in industry, scientific research, communications, medicine, the military, and the arts.

A -Industry

Powerful laser beams can be focused on a small spot to generate enormous temperatures. Consequently, the focused beams can readily and precisely heat, melt, or vaporize material. Lasers have been used, for example, to drill holes in diamonds, to shape machine tools, to trim microelectronics, to cut fashion patterns, to synthesize new material, and to attempt to induce controlled nuclear fusion. Highly directional laser beams are used for alignment in construction. Perfectly straight and uniformly sized tunnels, for example, may be dug using lasers for guidance. Powerful, short laser pulses also make high-speed photography with exposure times of only several trillionths of a second possible.

B -Scientific Research

Because laser light is highly directional and monochromatic, extremely small amounts of light scattering and small shifts in color caused by the interaction between laser light and matter can easily be detected. By measuring the scattering and color shifts, scientists can study molecular structures of matter. Chemical reactions can be selectively induced, and the existence of trace substances in samples can be detected. Lasers are also the most effective detectors of certain types of air pollution.

Scientists use lasers to make extremely accurate measurements. Lasers are used in this way for monitoring small movements associated with plate tectonics and for geographic surveys. Lasers have been used for precise determination (to within one inch) of the distance between Earth and the Moon, and in precise tests to confirm Einstein's theory of relativity. Scientists also have used lasers to determine the speed of light to an unprecedented accuracy. Very fast laser-activated switches are being developed for use in particle accelerators. Scientists also use lasers to trap single atoms and subatomic particles in order to study these tiny bits of matter

C -Communications

Laser light can travel a large distance in outer space with little reduction in signal strength. In addition, high-energy laser light can carry 1,000 times the television channels today carried by microwave signals. Lasers are therefore ideal for space communications. Low-loss optical fibers have been developed to transmit laser light for earthbound communication in telephone and computer systems. Laser techniques have also been used for high-density information recording. For instance, laser light simplifies the recording of a hologram, from which a three-dimensional image can be reconstructed with a laser beam. Lasers are also used to play audio CDs and videodiscs.

D -Medicine

Lasers have a wide range of medical uses. Intense, narrow beams of laser light can cut and cauterize certain body tissues in a small fraction of a second without damaging surrounding healthy tissues. Lasers have been used to ―weld‖ the retina, bore holes in the skull, vaporize lesions, and cauterize blood vessels. Laser surgery has virtually replaced older surgical procedures for eye disorders. Laser techniques have also been developed for lab tests of small biological samples.

E -Military Applications

Laser guidance systems for missiles, aircraft, and satellites have been constructed. Guns can be fitted with laser sights and range finders. The use of laser beams to destroy hostile ballistic missiles has been proposed, as in the Strategic Defense Initiative urged by U.S. president Ronald Reagan and the Ballistic Missile Defense program supported by President George W. Bush. The ability of tunable dye lasers to selectively excite an atom or molecule may open up more efficient ways to separate isotopes for construction of nuclear weapons.

V -LASER SAFETY

Because the eye focuses laser light just as it does other light, the chief danger in working with lasers is eye damage. Therefore, laser light should not be viewed either directly or reflected. Lasers sold and used commercially in the United States must comply with a strict set of laws enforced by the Center for Devices and Radiological Health (CDRH), a department of the Food and Drug Administration. The CDRH has divided lasers into six groups, depending on their power output, their emission duration, and the energy of the photons they emit. The classification is then attached to the laser as a sticker. The higher the laser‗s energy, the higher it‗s potential to injure. High-powered lasers of the Class IV type (the highest classification) generate a beam of energy that can start fires, burn flesh, and cause permanent eye damage whether the light is direct, reflected, or diffused. Canada uses the same classification system, and laser use in Canada is overseen by Health Canada‗s Radiation Protection Bureau.

Goggles blocking the specific color of photons that a laser produces are mandatory for the safe use of lasers. Even with goggles, direct exposure to laser light should be avoided

Light Absorption and Emission:

When a photon, or packet of light energy, is absorbed by an atom, the atom gains the energy of the photon, and one of the atom‗s electrons may jump to a higher energy level. The atom is then said to be excited. When an electron of an excited atom falls to a lower energy level, the atom may emit the electron‗s excess energy in the form of a photon. The energy levels, or orbital, of the atoms shown here have been greatly simplified to illustrate these absorption and emission processes. For a more accurate depiction of electron orbital, see the Atom article

Laser and Incandescent Light:

White light, such as that produced by an incandescent bulb, is composed of many colors of light—each with a different wavelength—and spreads out in all directions. Laser light consists of a single color (a single wavelength) and moves in one direction with the peaks and troughs of its waves in lockstep

31. MICROSCOPES

Microscopes give us a large image of a tiny object. The microscopes we use in school and at home trace their history back almost 400 years.

The first useful microscope was developed in the Netherlands between 1590 and 1608. There is almost as much confusion about the inventor as about the dates. Three different eyeglass makers have been given credit for the invention. The possible inventors are Hans Lippershey (who also developed the first real telescope), Hans Janssen, and his son, Zacharias.
Lens quality in early microscopes was often poor so the images were not very clear. But even these rather crude microscopes were a great help in learning more about animals and plants.

The microscope works a lot like a refracting telescope except that the object is very close to the objective lens. The clips on the microscope's flat stage hold the slide in place. A mirror at the bottom of the microscope reflects light rays up to the daphnia through a hole in the stage. Objective lenses magnify the image which is made even larger when we see it through the eyepiece lenses.

The objective lens is usually a compound lens, a combination of two lenses made from different kinds of glass. When only one lens is used, we often get distortion. This distortion (chromatic aberration) is caused because the colors making up light are not refracted (bent) the same amount when passing through a glass lens. When we use a compound lens, any distortion from the first lens is corrected by the second lens.

Different types of microscopes have been used to look at human cells, identify minerals, and solve crimes.
Microscopes are an essential tool in medicine too. They have been used to identify the causes of many deadly diseases like malaria and tuberculosis. Microscopes can also help to find out why a person or animal died.

Scientists can even use a microscope to figure out where illegal drugs come from. For example, looking at opium crystals through a microscope reveals different shapes depending on where the poppies they came from were grown. This information can help pinpoint the source of illegal drugs.

Compound Microscope:

Two convex lenses can form a microscope. The object lens is positioned close to the object to be viewed. It forms an upside-down and magnified image called a real image because the light rays actually pass through the place where the image lies. The ocular lens, or eyepiece lens, acts as a magnifying glass for this real image. The ocular lens makes the light rays spread more, so that they appear to come from a large inverted image beyond the object lens. Because light rays do not actually pass through this location, the image is called a virtual image

I -INTRODUCTION
Microscope, instrument used to obtain a magnified image of minute objects or minute details of objects.

II -OPTICAL MICROSCOPES
The most widely used microscopes are optical microscopes, which use visible light to create a magnified image of an object. The simplest optical microscope is the double-convex lens with a short focal length (see Optics). Double-convex lenses can magnify an object up to 15 times. The compound microscope uses two lenses, an objective lens and an ocular lens, mounted at opposite ends of a closed tube, to provide greater magnification than is possible with a single lens. The objective lens is composed of several lens elements that form an enlarged real image of the object being examined. The real image formed by the objective lens lies at the focal point of the ocular lens. Thus, the observer looking through the ocular lens sees an enlarged virtual image of the real image. The total magnification of a compound microscope is determined by the focal lengths of the two lens systems and can be more than 2000 times.

Optical microscopes have a firm stand with a flat stage to hold the material examined and some means for moving the microscope tube toward and away from the specimen to bring it into focus. Ordinarily, specimens are transparent and are mounted on slides—thin, rectangular pieces of clear glass that are placed on the stage for viewing. The stage has a small hole through which light can pass from a light source mounted underneath the stage—either a mirror that reflects natural light or a special electric light that directs light through the specimen.

In photomicrography, the process of taking photographs through a microscope, a camera is mounted directly above the microscope's eyepiece. Normally the camera does not contain a lens because the microscope itself acts as the lens system.

Microscopes used for research have a number of refinements to enable a complete study of the specimens. Because the image of a specimen is highly magnified and inverted, manipulating the specimen by hand is difficult. Therefore, the stages of high-powered research microscopes can by moved by micrometer screws, and in some microscopes, the stage can also be rotated. Research microscopes are also equipped with three or more objective lenses, mounted on a revolving head, so that the magnifying power of the microscope can be varied.

III -SPECIAL-PURPOSE OPTICAL MICROSCOPES
Different microscopes have been developed for specialized uses. The stereoscopic microscope, two low-powered microscopes arranged to converge on a single specimen, provides a three-dimensional image.
The petrographic microscope is used to analyze igneous and metamorphic rock. A Nicol prism or other polarizing device polarizes the light that passes through the specimen. Another Nicol prism or analyzer determines the polarization of the light after it has passed through the specimen. Rotating the stage causes changes in the polarization of light that can be measured and used to identify and estimate the mineral components of the rock.

The dark-field microscope employs a hollow, extremely intense cone of light concentrated on the specimen. The field of view of the objective lens lies in the hollow, dark portion of the cone and picks up only scattered light from the object. The clear portions of the specimen appear as a dark background, and the minute objects under study glow brightly against the dark field. This form of illumination is useful for transparent, unstained biological material and for minute objects that cannot be seen in normal illumination under the microscope.

The phase microscope also illuminates the specimen with a hollow cone of light. However, the cone of light is narrower and enters the field of view of the objective lens. Within the objective lens is a ring-shaped device that reduces the intensity of the light and introduces a phase shift of a quarter of a wavelength. This illumination causes minute variations of refractive index in a transparent specimen to become visible. This type of microscope is particularly effective for studying living tissue.

A typical optical microscope cannot resolve images smaller than the wavelength of light used to illuminate the specimen. An ultraviolet microscope uses the shorter wavelengths of the ultraviolet region of the light spectrum to increase resolution or to emphasize details by selective absorption (see Ultraviolet Radiation). Glass does not transmit the shorter wavelengths of ultraviolet light, so the optics in an ultraviolet microscope are usually quartz, fluorite, or aluminized-mirror systems. Ultraviolet radiation is invisible to human eyes, so the image must be made visible through phosphorescence (see Luminescence), photography, or electronic scanning.

The near-field microscope is an advanced optical microscope that is able to resolve details slightly smaller than the wavelength of visible light. This high resolution is achieved by passing a light beam through a tiny hole at a distance from the specimen of only about half the diameter of the hole. The light is played across the specimen until an entire image is obtained.

The magnifying power of a typical optical microscope is limited by the wavelengths of visible light. Details cannot be resolved that are smaller than these wavelengths. To overcome this limitation, the scanning interferometric aperture less microscope (SIAM) was developed. SIAM uses a silicon probe with a tip one nanometer (1 billionth of a meter) wide. This probe vibrates 200,000 times a second and scatters a portion of the light passing through an observed sample. The scattered light is then recombined with the unscattered light to produce an interference pattern that reveals minute details of the sample. The SIAM can currently resolve images 6500 times smaller than conventional light microscopes.

IV -ELECTRON MICROSCOPES
An electron microscope uses electrons to ―illuminate‖ an object. Electrons have a much smaller wavelength than light, so they can resolve much smaller structures. The smallest wavelength of visible light is about 4000 angstroms (40 millionths of a meter). The wavelength of electrons used in electron microscopes is usually about half an angstrom (50 trillionths of a meter).

Electron microscopes have an electron gun that emits electrons, which then strike the specimen. Conventional lenses used in optical microscopes to focus visible light do not work with electrons; instead, magnetic fields (see Magnetism) are used to create ―lenses‖ that direct and focus the electrons. Since electrons are easily scattered by air molecules, the interior of an electron microscope must be sealed at a very high vacuum. Electron microscopes also have systems that record or display the images produced by the electrons.

There are two types of electron microscopes: the transmission electron microscope (TEM), and the scanning electron microscope (SEM). In a TEM, the electron beam is directed onto the object to be magnified. Some of the electrons are absorbed or bounce off the specimen, while others pass through and form a magnified image of the specimen. The sample must be cut very thin to be used in a TEM, usually no more than a few thousand angstroms thick. A photographic plate or fluorescent screen beyond the sample records the magnified image. Transmission electron microscopes can magnify an object up to one million times. In a scanning electron microscope, a tightly focused electron beam moves over the entire sample to create a magnified image of the surface of the object in much the same way an electron beam scans an image onto the screen of a television. Electrons in the tightly focused beam might scatter directly off the sample or cause secondary electrons to be emitted from the surface of the sample. These scattered or secondary electrons are collected and counted by an electronic device. Each scanned point on the sample corresponds to a pixel on a television monitor; the more electrons the counting device detects, the brighter the pixel on the monitor is. As the electron beam scans over the entire sample, a complete image of the sample is displayed on the monitor.

An SEM scans the surface of the sample bit by bit, in contrast to a TEM, which looks at a relatively large area of the sample all at once. Samples scanned by an SEM do not need to be thinly sliced, as do TEM specimens, but they must be dehydrated to prevent the secondary electrons emitted from the specimen from being scattered by water molecules in the sample.
Scanning electron microscopes can magnify objects 100,000 times or more. SEMs are particularly useful because, unlike TEMs and powerful optical microscopes, they can produce detailed three-dimensional images of the surface of objects.

The scanning transmission electron microscope (STEM) combines elements of an SEM and a TEM and can resolve single atoms in a sample.
The electron probe microanalyzer, an electron microscope fitted with an X-ray spectrum analyzer, can examine the high-energy X rays emitted by the sample when it is bombarded with electrons. The identity of different atoms or molecules can be determined from their X-ray emissions, so the electron probe analyzer not only provides a magnified image of the sample, but also information about the sample's chemical composition.

V -SCANNING PROBE MICROSCOPES
A scanning probe microscope uses a probe to scan the surface of a sample and provides a three-dimensional image of atoms or molecules on the surface of the object. The probe is an extremely sharp metal point that can be as narrow as a single atom at the tip.

An important type of scanning probe microscope is the scanning tunneling microscope (STM). Invented in 1981, the STM uses a quantum physics phenomenon called tunneling to provide detailed images of substances that can conduct electricity. The probe is brought to within a few angstroms of the surface of the material being viewed, and a small voltage is applied between the surface and the probe. Because the probe is so close to the surface, electrons leak, or tunnel, across the gap between the probe and surface, generating a current. The strength of the tunneling current depends on the distance between the surface and the probe. If the probe moves closer to the surface, the tunneling current increases, and if the probe moves away from the surface, the tunneling current decreases. As the scanning mechanism moves along the surface of the substance, the mechanism constantly adjusts the height of the probe to keep the tunneling current constant. By tracking these minute adjustments with many scans back and forth along the surface, a computer can create a three-dimensional representation of the surface.

Another type of scanning probe microscope is the atomic force microscope (AFM). The AFM does not use a tunneling current, so the sample does not need to conduct electricity. As the metal probe in an AFM moves along the surface of a sample, the electrons in the probe are repelled by the electrons of the atoms in the sample and the AFM adjusts the height of the probe to keep the force on it constant. A sensing mechanism records the up-and-down movements of the probe and feeds the data into a computer, which creates a three-dimensional image of the surface of the sample.

32. COMPUTERS

Computer is an electronic device that can accept data, apply a series of logical processes to it and supply the results of these processes as information. Computers are also used to perform a complex series of mathematical calculations at very great speed which makes them great for the numerous purposes.

KINDS:

The two main kinds of computers are

1.Analog computer
2.Digital computer.

In analog computer the numbers are represented by magnitudes of physical quantities as voltage, magnitudes etc.

The digital computer is in which numbers are expressed directly as digits usually in the binary notion. The digital computers are however more useful and versatile.

COMPUTER BASICS:

Computer is mainly based on
1. Hardware
2. Software
3. Input

Hardware consists of devices, like the computer itself, the monitor, keyboard, printer, mouse and speakers. Inside your computer there are more bits of hardware, including the motherboard, where you would find the main processing chips that make up the central processing unit (CPU). The hardware processes the commands it receives from the software, and performs tasks or calculations.

Software is the name given to the programs that you install on the computer to perform certain types of activities

Input is when we type a command or click on an icon, we tell the computer what to do. That is called input.

WORKING:

A computer is based on various components which when combined together perform useful functions. There is a CPU, the central processing unit which performs all the computations. It is supported by memory which holds the current programmed and data and logic arrays which helps in the provision and moment of information around the system.
The program and data, text, figures and images or sounds are into in the computer which then processes the data and the outputs the results.

TYPES:

There are four main types of computers.
1. MICRO COMPUTERS
2. MINI COMPUTERS
3. MAIN FRAMES
4. SUPER COMPUTERS

Micro computers are the smallest and the most common and are used in small bussinesses, homes, schools.They are also referred as home computers.
The mini computers are also known as personal computers and are generally larger and used in medium sized businesses and university departments.
The mainframes are found in large organizations companies and government departments in advanced countries mostly.
The super computers are the most powerful of all as they are especially used for highly complex scientific tasks as analyzing results of nuclear physics experiments and weather forecasting.

MORE ABOUT COMPUTER:

Machine capable of executing instructions to perform operations on data. The distinguishing feature of a computer is its ability to store its own instructions. This ability makes it possible for a computer to perform many operations without the need for a person to enter new instructions each time. Modern computers are made of high-speed electronic components that enable the computer to perform thousands of operations each second.
Generations of computers are characterized by their technology. First-generation digital computers, developed mostly in the U.S. after World War II, used vacuum tubes and were enormous. The second generation, introduced c. 1960, used transistors and was the first successful commercial computers. Third-generation computers (late 1960s and 1970s) were characterized by miniaturization of components and use of integrated circuits. The microprocessor chip, introduced in 1974, defines fourth-generation computers.

Microprocessor:

A microprocessor is a computer processor on a microchip. It's sometimes called a logic chip. It is the "engine" that goes into motion when you turn your computer on. A microprocessor is designed to perform arithmetic and logic operations that make use of small number-holding areas called registers. Typical microprocessor operations include adding, subtracting, comparing two numbers, and fetching numbers from one area to another. These operations are the result of a set of instructions that are part of the microprocessor design. When the computer is turned on, the microprocessor is designed to get the first instruction from the basic input/output system (BIOS) that comes with the computer as part of its memory. After that, either the BIOS, or the operating system that BIOS loads into computer memory, or an application program is "driving" the microprocessor, giving it instructions to perform.

Digital Computers:

A digital computer is designed to process data in numerical form. Its circuits perform directly the mathematical operations of addition, subtraction, multiplication, and division. The numbers operated on by a digital computer are expressed in the binary system; binary digits, or bits, are 0 and 1, so that 0, 1, 10, 11, 100, 101, etc., correspond to 0, 1, 2, 3, 4, 5, etc. Binary digits are easily expressed in the computer circuitry by the presence (1) or absence (0) of a current or voltage. A series of eight consecutive bits is called a ―byte‖; the eight-bit byte permits 256 different ―on-off‖ combinations. Each byte can thus represent one of up to 256 alphanumeric characters, and such an arrangement is called a ―single-byte character set‖ (SBCS); the de facto standard for this representation is the extended ASCII character set. Some languages, such as Japanese, Chinese, and Korean, require more than 256 unique symbols. The use of two bytes, or 16 bits, for each symbol, however, permits the representation of up to 65,536 characters or ideographs. Such an arrangement is called a ―double-byte character set‖ (DBCS); Unicode is the international standard for such a character set. One or more bytes, depending on the computer's architecture, is sometimes called a digital word; it may specify not only the magnitude of the number in question, but also its sign (positive or negative), and may also contain redundant bits that allow automatic detection, and in some cases correction, of certain errors. A digital computer can store the results of its calculations for later use, can compare results with other data, and on the basis of such comparisons can change the series of operations it performs. Digital computers are used for reservations systems, scientific investigation, data-processing and word-processing applications, desktop publishing, electronic games, and many other purposes.

Analog Computers:

Computer in which continuously variable physical quantities, such as electrical potential, fluid pressure, or mechanical motion, are used to represent (analogously) the quantities in the problem to be solved. The analog system is set up according to initial conditions and then allowed to change freely. Answers to the problem are obtained by measuring the variables in the analog model. Analog computers are especially well suited to simulating dynamic systems; such simulations may be conducted in real time or at greatly accelerated rates, allowing experimentation by performing many runs with different variables. They have been widely used in simulating the operation of aircraft, nuclear power plants, and industrial chemical processes.

Minicomputers:

A minicomputer, a term no longer much used, is a computer of a size intermediate between a microcomputer and a mainframe. Typically, minicomputers have been stand-alone computers (computer systems with attached terminals and other devices) sold to small and mid-size businesses for general business applications and to large enterprises for department-level operations. In general, a minicomputer is a multiprocessing system capable of supporting from 4 to about 200 users simultaneously.

Microcomputers:

A digital computer whose central processing unit consists of a microprocessor, a single semiconductor integrated circuit chip. Once less powerful than larger computers, microcomputers are now as powerful as the minicomputers and super minicomputers of just several years ago. This is due in part to the growing processing power of each successive generation of microprocessor, plus the addition of mainframe computer features to the chip, such as floating-point mathematics, computation hardware, memory management, and multiprocessing support.

Microcomputers are the driving technology behind the growth of personal computers and workstations. The capabilities of today's microprocessors in combination with reduced power consumption have created a new category of microcomputers: hand-held devices. Some of these devices are actually general-purpose microcomputers: They have a liquid-crystal-display (LCD) screen and use an operating system that runs several general-purpose applications. Many others serve a fixed purpose, such as telephones that provide a display for receiving text-based pager messages and automobile navigation systems that use satellite-positioning signals to plot the vehicle's position.

Mainframe:

A mainframe (also known as "big iron") is a high-performance computer used for large-scale computing purposes that require greater availability and security than a smaller-scale machine can offer. Historically, mainframes have been associated with centralized rather than distributed computing, although that distinction is blurring as smaller computers become more powerful and mainframes become more multi-purpose.

A mainframe may support 100-500 users at one time. Typically, mainframes have a word length of 64 bits and are significantly faster and have greater capacity than the minicomputer and the microcomputer.

Supercomputers:

Supercomputer is a computer that performs at or near the currently highest operational rate for computers. A supercomputer is typically used for scientific and engineering applications that must handle very large databases or do a great amount of computation (or both). At any given time, there are usually a few well-publicized supercomputers that operate at the very latest and always incredible speeds. The term is also sometimes applied to far slower (but still impressively fast) computers. Most supercomputers are really multiple computers that perform parallel processing. In general, there are two parallel processing approaches: symmetric multiprocessing (SMP) and massively parallel processing (MPP).

Hardware:

Mechanical and electronic parts that constitute a computer system, as distinguished from the computer programs (Software) that drive the system. The main hardware elements are the Central Processing Unit, Disk or magnetic tape data storage devices, Cathode-Ray Tube display terminals, keyboards, and Printers. In operation, a computer is both hardware and software. One is useless without the other. The hardware design specifies the commands it can follow, and the software instructions tell it what to do.

Software:

A set of instructions that cause a computer to perform one or more tasks. The set of instructions is often called a program or, if the set is particularly large and complex, a system. Computers cannot do any useful work without instructions from software; thus a combination of software and hardware (the computer) is necessary to do any computerized work. A program must tell the computer each of a set of minuscule tasks to perform, in a framework of logic, such that the computer knows exactly what to do and when to do it.

Input Devices:

An input device is a hardware mechanism that transforms information in the external world for consumption by a computer. Often, input devices are under direct control by a human user, who uses them to communicate commands or other information to be processed by the computer, which may then transmit feedback to the user through an output device. Input and output devices together make up the hardware interface between a computer and the user or external world. Typical examples of input devices include keyboards and mice. However, there are others which provide many more degrees of freedom. In general, any sensor which monitors, scans for and accepts information from the external world can be considered an input device, whether or not the information is under the direct control of a user.

Keyboard:

In computing, a keyboard is a peripheral partially modeled after the typewriter keyboard. Keyboards are designed to input text and characters, as well as to operate a computer. Physically, keyboards are an arrangement of rectangular buttons, or "keys". Keyboards typically have characters engraved or printed on the keys; in most cases, each press of a key corresponds to a single written symbol. However, to produce some symbols requires pressing and holding several keys simultaneously or in sequence; other keys do not produce any symbol, but instead affect the operation of the computer or the keyboard itself.
Roughly 50% of all keyboard keys produce letters, numbers or signs (characters). Other keys can produce actions when pressed, and other actions are available by the simultaneous pressing of more than one action key.

Mouse:

A device that controls the movement of the cursor or pointer on a display screen. A mouse is a small object you can roll along a hard, flat surface. Its name is derived from its shape, which looks a bit like a mouse, its connecting wire that one can imagine to be the mouse's tail, and the fact that one must make it scurry along a surface. As you move the mouse, the pointer on the display screen moves in the same direction. Mice contain at least one button and sometimes as many as three, which have different functions depending on what program is running.

Output Devices:

Any machine capable of representing information from a computer. This includes display screens, printers, plotters, and synthesizers.

Display Screen:

The monitor displays the video and graphics information generated by the computer through the video card. Monitors are very similar to televisions but display information at a much higher quality. The Monitor is also known as monitor. The term monitor, however, usually refers to the entire box, whereas display screen can mean just the screen.

Printer:

A printer outputs data that is seen on the computer screen. Most printers are used through a parallel port, but some newer ones use USB connections. USB is somewhat faster, but there's not much of a difference for printers. Networked computers usually print to a printer through the network card. The most crucial printer measurement is its dots per inch rating. Although this can be misleading, a higher number is generally better. Printers are best chosen by actually seeing the quality of the printer output.

Scanner:

A scanner is a piece of hardware used to scan a document, i.e., create a digital copy. Although flatbed scanners are the most common type and operate much like a photocopy machine, there are many types of scanners, including some that never touch the document itself. Scanners use a variety of connection formats including Parallel Port, USB, and SCSI. USB is simple, SCSI is fast, and Parallel Port is extremely slow.

CPU (Central Processing Unit)

Stands for "Central Processing Unit." This is the pretty much the brain of computer. It processes everything from basic instructions to complex functions. Any time something needs to be computed, it gets sent to the CPU.

Generally, the CPU is a single microchip, but that doesn't necessarily have to be the case. In the consumer desktop and laptop market, the CPU market is dominated by Intel, AMD, and IBM. These manufacturers supply the computer makers such as Dell, HP, and Apple.

Due to its importance to every computing task, the speed of the CPU, usually measured in gigahertz (GHz) is the number that most vendors use in their marketing campaigns. In the past, the larger the number, the faster the computer could be expected to be. However, in recent years, the speed of the CPU has had less impact as other components of a computer take on more and more of the workload. Also, differences in technology mean that a slower chip that performs more calculations per cycle can actually be faster than a higher rate chip doing fewer calculations per cycle.

Bit:

A binary digit. The term was first used in 1946 by John Tukey, a leading statistician and adviser to five presidents. In the computer, electronics, and communications fields, ―bit‖ is generally understood as a shortened form of ―binary digit.‖ In a numerical binary system, a bit is either a 0 or 1. Bits are generally used to indicate situations that can take one of two values or one of two states, for example, on and off, true or false, or yes or no. If, by convention, 1 represents a particular state, then 0 represents the other state. For example, if 1 stands for ―yes,‖ then 0 stands for ―no.‖ A bit is abbreviated with a small "b".

Byte:

The common unit of computer storage from desktop computer to mainframe. The term byte was coined by Dr. Werner Buchholz in July 1956, during the early design phase for the IBM Stretch computer. It is made up of eight binary digits (bits). A ninth bit may be used in the memory circuits as a parity bit for error checking. The term was originally coined to mean the smallest addressable group of bits in a computer, which has not always been eight. A byte is abbreviated with a "B".

RAM:

RAM stands for Random Access Memory. Computer main memory in which specific contents can be accessed (read or written) directly by the CPU in a very short time regardless of the sequence (and hence location) in which they were recorded. Two types of memory are possible with random-access circuits, static RAM (SRAM) and dynamic RAM (DRAM). A single memory chip is made up of several million memory cells. In a SRAM chip, each memory cell stores a binary digit (1 or 0) for as long as power is supplied. In a DRAM chip, the charge on individual memory cells must be refreshed periodically in order to retain data. Because it has fewer components, DRAM requires less chip area than SRAM; hence a DRAM chip can hold more memory, though its access time is slower. The size of the RAM (measured by kilobytes) is an important indicator of the capacity of the computer.

ROM:

ROM stands for Read Only Memory. A memory chip that permanently stores instructions and data. Also known as "mask ROM," its content is created in the last masking stage of the chip manufacturing process, and it cannot be changed. Once data has been written onto a ROM chip, it cannot be removed and can only be read.
Unlike main memory (RAM), ROM retains its contents even when the computer is turned off. ROM is referred to as being nonvolatile, whereas RAM is volatile.

Computer Networking:

A computer network is an interconnected group of computers. Networks may be classified by the network layer at which they operate according to basic reference models considered as standards in the industry, such as the four-layer Internet Protocol Suite model. While the seven-layer Open Systems Interconnection (OSI) reference model is better known in academia, the majority of networks use the Internet Protocol Suite (IP).
Computer networks may be classified according to the scale.

Personal Area Network (PAN)

A personal area network (PAN) is the interconnection of information technology devices within the range of an individual person, typically within a range of 10 meters. For example, a person traveling with a laptop, a personal digital assistant (PDA), and a portable printer could interconnect them without having to plug anything in, using some form of wireless technology. Typically, this kind of personal area network could also be interconnected without wires to the Internet or other networks.

Local Area Network (LAN)

Communications network connecting computers by wire, cable, or fiber optics link. Usually serves parts of an organization located close to one another, generally in the same building or within 2 miles of one another. Allows users to share software, hardware and data. The first LAN put into service occurred in 1964 at the Livermore Laboratory to support atomic weapons research. LANs spread to the public sector in the late 1970s and were used to create high-speed links between several large central computers at one site.

Initially, LANs were limited to a range of 185 meters or 600 feet and could not include more than 30 computers. Today, a LAN could connect a max of 1024 computers at a max distance of 900 meters or 2700 feet.

Campus Area Network (CAN)

A campus area network (CAN) is a computer network interconnecting a few local area networks (LANs) within a university campus or corporate campus. Campus area network may link a variety of campus buildings including departments, the university library and student halls of residence. A campus area network is larger than a local area network but smaller than a metropolitan area network (MAN) or wide area network (WAN). CAN can also stand for corporate area network.

Metropolitan area network (MAN)

A metropolitan area network (MAN) is a network that interconnects users with computer resources in a geographic area or region larger than that covered by even a large local area network (LAN) but smaller than the area covered by a wide area network (WAN). The term is applied to the interconnection of networks in a city into a single larger network (which may then also offer efficient connection to a wide area network). It is also used to mean the interconnection of several local area networks by bridging them with backbone lines. The latter usage is also sometimes referred to as a campus network.

MAN networks use a different standard for communications; 802.6 as assigned by the Institute of Electrical and Electronics Engineers (IEEE), which uses a different bus technology to transmit and receive data than most larger or smaller networks. This allows MAN networks to operate more efficiently than they might if they were simply LAN networks linked together.

Wide area network (WAN)

The wide area network, often referred to as a WAN, is a communications network that makes use of existing technology to connect local computer networks into a larger working network that may cover both national and international locations. This is in contrast to both the local area network and the metropolitan area network, which provides communication within a restricted geographic area. The largest WAN in existence is the Internet.

Arithmetic Logic Unit (ALU)

In computing, an arithmetic logic unit (ALU) is a digital circuit that performs arithmetic and logical operations. The ALU is a fundamental building block of the central processing unit of a computer, and even the simplest microprocessors contain one for purposes such as maintaining timers. The processors found inside modern CPUs and GPU have inside them very powerful and very complex ALU; a single component may contain a number of ALU.

Mathematician John von Neumann proposed the ALU concept in 1945, when he wrote a report on the foundations for a new computer called the EDVAC.

Control Unit:

The control unit is the circuitry that controls the flow of information through the processor, and coordinates the activities of the other units within it. In a way, it is the "brain within the brain", as it controls what happens inside the processor, which in turn controls the rest of the PC.

The functions performed by the control unit vary greatly by the internal architecture of the CPU, since the control unit really implements this architecture. On a regular processor that executes x86 instructions natively, the control unit performs the tasks of fetching, decoding, managing execution and then storing results. On a processor with a RISC core the control unit has significantly more work to do. It manages the translation of x86 instructions to RISC micro-instructions, manages scheduling the micro-instructions between the various execution units, and juggles the output from these units to make sure they end up where they are supposed to go. On one of these processors the control unit may be broken into other units (such as a scheduling unit to handle scheduling and a retirement unit to deal with results coming from the pipeline) due to the complexity of the job it must perform.

Modem:

Equipment that converts digital signals into analog signals for purpose of transmission over a telephone line. Signal is then converted back to digital form so that it can be processed by a receiving computer. Modems are typically used to link computers via telephone lines. Short for modulator-demodulator.
The speed at which a modem transmits data is measured in units called bits per second or bps. The first modems ran at even less than 300 bps. Now 1200, 2400, and 9600 bps modems are considered slow. The faster models reach speeds of 14,400 and 28,800 bps. The faster the modem, the faster the data (for example, images from the Web) appear. Even a 28,800 bps modem, however, cannot compare to the several million bps speed that a campus Ethernet connection gives you.

Register:

A small, high-speed computer circuit that holds values of internal operations, such as the address of the instruction being executed and the data being processed. When a program is debugged, register contents may be analyzed to determine the computer's status at the time of failure.
In microcomputer assembly language programming, programmers look at the contents of registers routinely. Assembly languages in larger computers are often at a higher level.

Cache Memory:

Cache memory is extremely fast memory that is built into a computer‛s central processing unit (CPU), or located next to it on a separate chip. The CPU uses cache memory to store instructions that are repeatedly required to run programs, improving overall system speed. The advantage of cache memory is that the CPU does not have to use the motherboard‛s system bus for data transfer. Whenever data must be passed through the system bus, the data transfer speed slows to the motherboard‛s capability. The CPU can process data much faster by avoiding the bottleneck created by the system bus.

Cache that is built into the CPU is faster than separate cache, running at the speed of the microprocessor itself. However, separate cache is still roughly twice as fast as Random Access Memory (RAM). Cache is more expensive than RAM, but it is well worth getting a CPU and motherboard with built-in cache in order to maximize system performance.

Computer Virus:

A virus is a program designed to infect and potentially damage files on a computer that receives it. The code for a virus is hidden within an existing program—such as a word processing or spreadsheet program—and when that program is launched, the virus inserts copies of itself into other programs on the system to infect them as well. Because of this ability to reproduce itself, a virus can quickly spread to other programs, including the computer's operating system. A virus may be resident on a system for a period of time before taking any action detectable to the user. The impact of other viruses may be felt immediately. Some viruses cause little or no damage. For example, a virus may manifest itself as nothing more than a message that appears on the screen at certain intervals. Other viruses are much more destructive and can result in lost or corrupted files and data. At their worst, viruses may render a computer unusable, necessitating the reinstallation of the operating system and applications.

33. SATELLITES

Satellite technology has emerged tremendously over the last 50 years since Arthur C. Clarke first invented it. Today, satellite technology is all around us and has become a very useful, everyday application of modern telecommunications. Satellite systems can provide a variety of services including broadband communication systems, satellite-based video, audio, internet and data distribution networks, as well as worldwide customer service and support.

What is a satellite?

An artificial satellite is a manmade object placed into orbit around the Earth for the purpose of scientific research, weather reports, or military reconnaissance. Scientific satellites are set into orbit to observe the space environment, the Earth, the Sun, stars and extra galactic objects. These satellites have retrieved a huge amount of information helpful to scientific research. Weather satellites are used every day for meteorological forecasts and in shipping. Also military satellites play an important role in today's modern military. Satellites are extremely important today. All artificial satellites have certain features in common. They all include radar systems, sensors like optical devices in observation satellites and receivers and transmitters in communication satellites. Solar cells are used to generate power for the satellites and in some cases, nuclear power is used. All satellites need altitude-control equipment to keep the satellite in the desired orbit.

Orbit of a Satellite:

The orbit of the satellite is achieved when it is given a horizontal velocity of 17,500 mph at sea level causing the Earth's surface to curve away and as fast as it curves away gravity pulls the object downward and at this point the satellite achieved orbit. As the altitude of the satellite increases, its velocity decreases and its period increases. The period of satellite is the time the satellite takes to make on revolution around the Earth. Satellites in later orbit are called synchronous satellites. If the satellite orbits in a equatorial plane, it is called geostationary which means it is always over the same place on earth at all times. This form of orbit is used in weather for reports of a certain area at all times. The orbit of a satellite is very scientific but not hard to understand.

34. ANTIBIOTICS

A chemical substance derivable from a mold or bacterium that kills microorganisms and cures infections.

Antibiotics are drugs used to kill or harm specific bacteria. Since their discovery in the 1930s, antibiotics have made it possible to cure diseases caused by bacteria such as pneumonia, tuberculosis, and meningitis - saving the lives of millions of people around the world.

But antibiotics must be used wisely. Because bacteria are living organisms, they are always changing in an effort to resist the drugs that can kill them. When antibiotics are used incorrectly, bacteria can adapt and become resistant. Antibiotics are then no longer useful in fighting them. Antibiotic resistance is now a major public health issue. The correct use of these drugs is the best way to ensure that antibiotics remain useful in treating infections.

I -INTRODUCTION

Antibiotics (Greek anti, ―against‖; bios, ―life‖) are chemical compounds used to kill or inhibit the growth of infectious organisms. Originally the term antibiotic referred only to organic compounds, produced by bacteria or molds, that are toxic to other microorganisms. The term is now used loosely to include synthetic and semi synthetic organic compounds. Antibiotic refers generally to antibacterial; however, because the term is loosely defined, it is preferable to specify compounds as being antimalarials, antiviral, or antiprotozoals. All antibiotics share the property of selective toxicity: They are more toxic to an invading organism than they are to an animal or human host. Penicillin is the most well-known antibiotic and has been used to fight many infectious diseases, including syphilis, gonorrhea, tetanus, and scarlet fever. Another antibiotic, streptomycin, has been used to combat tuberculosis.

II -HISTORY

Although the mechanisms of antibiotic action were not scientifically understood until the late 20th century, the principle of using organic compounds to fight infection has been known since ancient times. Crude plant extracts were used medicinally for centuries, and there is anecdotal evidence for the use of cheese molds for topical treatment of infection. The first observation of what would now be called an antibiotic effect was made in the 19th century by French chemist Louis Pasteur, who discovered that certain saprophytic bacteria can kill anthrax bacilli.

In the first decade of the 20th century, German physician and chemist Paul Ehrlich began experimenting with the synthesis of organic compounds that would selectively attack an infecting organism without harming the host organism. His experiments led to the development, in 1909, of salvarsan, a synthetic compound containing arsenic, which exhibited selective action against spirochetes, the bacteria that cause syphilis. Salvarsan remained the only effective treatment for syphilis until the purification of penicillin in the 1940s. In the 1920s British bacteriologist Sir Alexander Fleming, who later discovered penicillin, found a substance called lysozyme in many bodily secretions, such as tears and sweat, and in certain other plant and animal substances. Lysozyme has some antimicrobial activity, but it is not clinically useful.

Penicillin, the archetype of antibiotics, is a derivative of the mold Penicillium notatum. Penicillin was discovered accidentally in 1928 by Fleming, who showed its effectiveness in laboratory cultures against many disease-producing bacteria. This discovery marked the beginning of the development of antibacterial compounds produced by living organisms. Penicillin in its original form could not be given by mouth because it was destroyed in the digestive tract and the preparations had too many impurities for injection. No progress was made until the outbreak of World War II stimulated renewed research and the Australian pathologist Sir Howard Florey and German-British biochemist Ernst Chain purified enough of the drug to show that it would protect mice from infection. Florey and Chain then used the purified penicillin on a human patient who had staphylococcal and streptococcal septicemia with multiple abscesses and osteomyelitis. The patient, gravely ill and near death, was given intravenous injections of a partly purified preparation of penicillin every three hours. Because so little was available, the patient's urine was collected each day; the penicillin was extracted from the urine and used again. After five days the patient's condition improved vastly. However, with each passage through the body, some penicillin was lost. Eventually the supply ran out and the patient died.

The first antibiotic to be used successfully in the treatment of human disease was tyrothricin, isolated from certain soil bacteria by American bacteriologist Rene Dubos in 1939. This substance is too toxic for general use, but it is employed in the external treatment of certain infections. Other antibiotics produced by a group of soil bacteria called actinomycetes have proved more successful. One of these, streptomycin, discovered in 1944 by American biologist Selman Waksman and his associates, was, in its time, the major treatment for tuberculosis.

Since antibiotics came into general use in the 1950s, they have transformed the patterns of disease and death. Many diseases that once headed the mortality tables—such as tuberculosis, pneumonia, and septicemia—now hold lower positions. Surgical procedures, too, have been improved enormously, because lengthy and complex operations can now be carried out without a prohibitively high risk of infection. Chemotherapy has also been used in the treatment or prevention of protozoal and fungal diseases, especially malaria, a major killer in economically developing nations (see Third World). Slow progress is being made in the chemotherapeutic treatment of viral diseases. New drugs have been developed and used to treat shingles (see herpes) and chicken pox. There is also a continuing effort to find a cure for acquired immunodeficiency syndrome (AIDS), caused by the human immunodeficiency virus (HIV).

III -CLASSIFICATION

Antibiotics can be classified in several ways. The most common method classifies them according to their action against the infecting organism. Some antibiotics attack the cell wall; some disrupt the cell membrane; and the majority inhibit the synthesis of nucleic acids and proteins, the polymers that make up the bacterial cell. Another method classifies antibiotics according to which bacterial strains they affect: staphylococcus, streptococcus, or Escherichia coli, for example. Antibiotics are also classified on the basis of chemical structure, as penicillins, cephalosporins, aminoglycosides, tetracyclines, macrolides, or sulfonamides, among others.

A -Mechanisms of Action

Most antibiotics act by selectively interfering with the synthesis of one of the large-molecule constituents of the cell—the cell wall or proteins or nucleic acids. Some, however, act by disrupting the cell membrane (see Cell Death and Growth Suppression below). Some important and clinically useful drugs interfere with the synthesis of peptidoglycan, the most important component of the cell wall. These drugs include the B-lactam antibiotics, which are classified according to chemical structure into penicillins, cephalosporins, and carbapenems. All these antibiotics contain a B-lactam ring as a critical part of their chemical structure, and they inhibit synthesis of peptidoglycan, an essential part of the cell wall. They do not interfere with the synthesis of other intracellular components. The continuing buildup of materials inside the cell exerts ever-greater pressure on the membrane, which is no longer properly supported by peptidoglycan. The membrane gives way, the cell contents leak out, and the bacterium dies. These antibiotics do not affect human cells because human cells do not have cell walls.

Many antibiotics operate by inhibiting the synthesis of various intracellular bacterial molecules, including DNA, RNA, ribosomes, and proteins. The synthetic sulfonamides are among the antibiotics that indirectly interfere with nucleic acid synthesis. Nucleic-acid synthesis can also be stopped by antibiotics that inhibit the enzymes that assemble these polymers—for example, DNA polymerase or RNA polymerase. Examples of such antibiotics are actinomycin, rifamicin, and rifampicin, the last two being particularly valuable in the treatment of tuberculosis. The quinolone antibiotics inhibit synthesis of an enzyme responsible for the coiling and uncoiling of the chromosome, a process necessary for DNA replication and for transcription to messenger RNA. Some antibacterials affect the assembly of messenger RNA, thus causing its genetic message to be garbled. When these faulty messages are translated, the protein products are nonfunctional. There are also other mechanisms: The tetracyclines compete with incoming transfer-RNA molecules; the aminoglycosides cause the genetic message to be misread and a defective protein to be produced; chloramphenicol prevents the linking of amino acids to the growing protein; and puromycin causes the protein chain to terminate prematurely, releasing an incomplete protein.

B -Range of Effectiveness

In some species of bacteria the cell wall consists primarily of a thick layer of peptidoglycan. Other species have a much thinner layer of peptidoglycan and an outer as well as an inner membrane. When bacteria are subjected to Gram's stain, these differences in structure affect the differential staining of the bacteria with a dye called gentian violet. The differences in staining coloration (gram-positive bacteria appear purple and gram-negative bacteria appear colorless or reddish, depending on the process used) are the basis of the classification of bacteria into gram-positive (those with thick peptidoglycan) and gram-negative (those with thin peptidoglycan and an outer membrane), because the staining properties correlate with many other bacterial properties. Antibacterials can be further subdivided into narrow-spectrum and broad-spectrum agents. The narrow-spectrum penicillins act against many gram-positive bacteria. Aminoglycosides, also narrow-spectrum, act against many gram-negative as well as some gram-positive bacteria. The tetracyclines and chloramphenicols are both broad-spectrum drugs because they are effective against both gram-positive and gram-negative bacteria.

C -Cell Death and Growth Suppression

Antibiotics may also be classed as bactericidal (killing bacteria) or bacteriostatic (stopping bacterial growth and multiplication). Bacteriostatic drugs are nonetheless effective because bacteria that are prevented from growing will die off after a time or be killed by the defense mechanisms of the host. The tetracyclines and the sulfonamides are among the bacteriostatic antiobiotics. Antibiotics that damage the cell membrane cause the cell's metabolites to leak out, thus killing the organism. Such compounds, including penicillins and cephalosporins, are therefore classed as bactericidal.

IV -TYPES OF ANTIBIOTICS

Following is a list of some of the more common antibiotics and examples of some of their clinical uses. This section does not include all antibiotics nor all of their clinical applications.

A -Penicillins

Penicillins are bactericidal, inhibiting formation of the cell wall. There are four types of penicillins: the narrow-spectrum penicillin-G types, ampicillin and its relatives, the penicillinase-resistants, and the extended spectrum penicillins that are active against pseudomonas. Penicillin-G types are effective against gram-positive strains of streptococci, staphylococci, and some gram-negative bacteria such as meningococcus. Penicillin-G is used to treat such diseases as syphilis, gonorrhea, meningitis, anthrax, and yaws. The related penicillin V has a similar range of action but is less effective. Ampicillin and amoxicillin have a range of effectiveness similar to that of penicillin-G, with a slightly broader spectrum, including some gram-negative bacteria. The penicillinase-resistants are penicillins that combat bacteria that have developed resistance to penicillin-G. The antipseudomonal penicillins are used against infections caused by gram-negative Pseudomonas bacteria, a particular problem in hospitals. They may be administered as a prophylactic in patients with compromised immune systems, who are at risk from gram-negative infections.

Side effects of the penicillins, while relatively rare, can include immediate and delayed allergic reactions—specifically, skin rashes, fever, and anaphylactic shock, which can be fatal.

B -Cephalosporin

Like the penicillins, cephalosporins have a B-lactam ring structure that interferes with synthesis of the bacterial cell wall and so are bactericidal. Cephalosporins are more effective than penicillin against gram-negative bacilli and equally effective against gram-positive cocci. Cephalosporins may be used to treat strains of meningitis and as a prophylactic for orthopedic, abdominal, and pelvic surgery. Rare hypersensitive reactions from the cephalosporins include skin rash and, less frequently, anaphylactic shock.

C -Aminoglycosides

Streptomycin is the oldest of the aminoglycosides. The aminoglycosides inhibit bacterial protein synthesis in many gram-negative and some gram-positive organisms. They are sometimes used in combination with penicillin. The members of this group tend to be more toxic than other antibiotics. Rare adverse effects associated with prolonged use of aminoglycosides include damage to the vestibular region of the ear, hearing loss, and kidney damage.

D -Tetracyclines

Tetracyclines are bacteriostatic, inhibiting bacterial protein synthesis. They are broad-spectrum antibiotics effective against strains of streptococci, gram-negative bacilli, rickettsia (the bacteria that causes typhoid fever), and spirochetes (the bacteria that causes syphilis). They are also used to treat urinary-tract infections and bronchitis. Because of their wide range of effectiveness, tetracyclines can sometimes upset the balance of resident bacteria that are normally held in check by the body's immune system, leading to secondary infections in the gastrointestinal tract and vagina, for example. Tetracycline use is now limited because of the increase of resistant bacterial strains.

E -Macrolides

The macrolides are bacteriostatic, binding with bacterial ribosomes to inhibit protein synthesis. Erythromycin, one of the macrolides, is effective against gram-positive cocci and is often used as a substitute for penicillin against streptococcal and pneumococcal infections. Other uses for macrolides include diphtheria and bacteremia. Side effects may include nausea, vomiting, and diarrhea; infrequently, there may be temporary auditory impairment.

F -Sulfonamides
The sulfonamides are synthetic bacteriostatic, broad-spectrum antibiotics, effective against most gram-positive and many gram-negative bacteria. However, because many gram-negative bacteria have developed resistance to the sulfonamides, these antibiotics are now used only in very specific situations, including treatment of urinary-tract infection, against meningococcal strains, and as a prophylactic for rheumatic fever. Side effects may include disruption of the gastrointestinal tract and hypersensitivity.

V -PRODUCTION
The production of a new antibiotic is lengthy and costly. First, the organism that makes the antibiotic must be identified and the antibiotic tested against a wide variety of bacterial species. Then the organism must be grown on a scale large enough to allow the purification and chemical analysis of the antibiotic and to demonstrate that it is unique. This is a complex procedure because there are several thousand compounds with antibiotic activity that have already been discovered, and these compounds are repeatedly rediscovered. After the antibiotic has been shown to be useful in the treatment of infections in animals, larger-scale preparation can be undertaken.

Commercial development requires a high yield and an economic method of purification. Extensive research may be needed to increase the yield by selecting improved strains of the organism or by changing the growth medium. The organism is then grown in large steel vats, in submerged cultures with forced aeration. The naturally fermented product may be modified chemically to produce a semisynthetic antibiotic. After purification, the effect of the antibiotic on the normal function of host tissues and organs (its pharmacology), as well as its possible toxic actions (toxicology), must be tested on a large number of animals of several species. In addition, the effective forms of administration must be determined. Antibiotics may be topical, applied to the surface of the skin, eye, or ear in the form of ointments or creams. They may be oral, or given by mouth, and either allowed to dissolve in the mouth or swallowed, in which case they are absorbed into the bloodstream through the intestines. Antibiotics may also be parenteral, or injected intramuscularly, intravenously, or subcutaneously; antibiotics are administered parenterally when fast absorption is required.

In the United States, once these steps have been completed, the manufacturer may file an Investigational New Drug Application with the Food and Drug Administration (FDA). If approved, the antibiotic can be tested on volunteers for toxicity, tolerance, absorption, and excretion. If subsequent tests on small numbers of patients are successful, the drug can be used on a larger group, usually in the hundreds. Finally a New Drug Application can be filed with the FDA, and, if this application is approved, the drug can be used generally in clinical medicine. These procedures, from the time the antibiotic is discovered in the laboratory until it undergoes clinical trial, usually extend over several years.

VI -RISKS AND LIMITATIONS
The use of antibiotics is limited because bacteria have evolved defenses against certain antibiotics. One of the main mechanisms of defense is inactivation of the antibiotic. This is the usual defense against penicillins and chloramphenicol, among others. Another form of defense involves a mutation that changes the bacterial enzyme affected by the drug in such a way that the antibiotic can no longer inhibit it. This is the main mechanism of resistance to the compounds that inhibit protein synthesis, such as the tetracyclines.

All these forms of resistance are transmitted genetically by the bacterium to its progeny. Genes that carry resistance can also be transmitted from one bacterium to another by means of plasmids, chromosomal fragments that contain only a few genes, including the resistance gene. Some bacteria conjugate with others of the same species, forming temporary links during which the plasmids are passed from one to another. If two plasmids carrying resistance genes to different antibiotics are transferred to the same bacterium, their resistance genes can be assembled onto a single plasmid. The combined resistances can then be transmitted to another bacterium, where they may be combined with yet another type of resistance. In this way, plasmids are generated that carry resistance to several different classes of antibiotic. In addition, plasmids have evolved that can be transmitted from one species of bacteria to another, and these can transfer multiple antibiotic resistance between very dissimilar species of bacteria.
The problem of resistance has been exacerbated by the use of antibiotics as prophylactics, intended to prevent infection before it occurs. Indiscriminate and inappropriate use of antibiotics for the treatment of the common cold and other common viral infections, against which they have no effect, removes antibiotic-sensitive bacteria and allows the development of antibiotic-resistant bacteria. Similarly, the use of antibiotics in poultry and livestock feed has promoted the spread of drug resistance and has led to the widespread contamination of meat and poultry by drug-resistant bacteria such as Salmonella.

In the 1970s, tuberculosis seemed to have been nearly eradicated in the developed countries, although it was still prevalent in developing countries. Now its incidence is increasing, partly due to resistance of the tubercle bacillus to antibiotics. Some bacteria, particularly strains of staphylococci, are resistant to so many classes of antibiotics that the infections they cause are almost untreatable. When such a strain invades a surgical ward in a hospital, it is sometimes necessary to close the ward altogether for a time. Similarly, plasmodia, the causative organisms of malaria, have developed resistance to antibiotics, while, at the same time, the mosquitoes that carry plasmodia have become resistant to the insecticides that were once used to control them. Consequently, although malaria had been almost entirely eliminated, it is now again rampant in Africa, the Middle East, Southeast Asia, and parts of Latin America. Furthermore, the discovery of new antibiotics is now much less common than in the past

35. VACCINES

Immunogenic consisting of a suspension of weakened or dead pathogenic cells injected in order to stimulate the production of antibodies can be defined as vaccines.

How Vaccines Work

Disease causing organisms have at least two distinct effects on the body The first effect is exhibiting symptoms such as fever, nausea, vomiting, diarrhea, rash, and many others. the second effect generally leads to eventual recovery from the infection: the disease causing organism induces an immune response in the infected host. As the response increases in strength over time, the infectious agents are slowly reduced in number until symptoms disappear and recovery is complete.

The disease causing organisms contain proteins called "antigens" which stimulate the immune response. The resulting immune response is multi-fold and includes the synthesis of proteins called "antibodies." These proteins bind to the disease causing organisms and lead to their eventual destruction. In addition, "memory cells" are produced in an immune response. These are cells which remain in the blood stream, sometimes for the life span of the host, ready to mount a quick protective immune response against subsequent infections with the particular disease causing agent which induced their production. If such an infection were to occur, the memory cells would respond so quickly that the resulting immune response could inactivate the disease causing agents, and symptoms would be prevented. This response is often so rapid that infection doesn't develop - and we get immune from infection.

Vaccines are effective in preventing disease not only in individuals, but also in communities. This type of protection is called "herd immunity." When a disease spreads from one human to another, it requires both an infected individual to spread it and a susceptible individual to catch it. Herd immunity works by decreasing the numbers of susceptible people. When the number of susceptible people drops low enough, the disease will disappear from the community because there are not enough people to carry on the catch-and-infect cycle. The greater the proportion of vaccinated members of the community, the more rapidly the disease will disappear.

36. FERTILIZERS

Any substance such as manure or a mixture of nitrates used to make soil more fertile are fertilizers.

Fertilizers are plant nutrients. Nutrients exist naturally in the earth's soil and atmosphere, and in animal manure. However, naturally occurring nutrients are not always available in the forms that plants can use. Therefore, man-made fertilizer is vital to food production. Man-made and natural fertilizers contain the same ingredients, but man-made fertilizers act more quickly and are less susceptible to weather changes. Farmers, ranchers and gardeners add these fertilizers directly to the soil, where they can be absorbed by plants for healthy growth. Incorporated into a program of best management practices, which includes soil testing, man-made fertilizer use leads to higher crop yields and greater environmental protection.

Fertilizer, natural or synthetic chemical substance or mixture used to enrich soil so as to promote plant growth. Plants do not require complex chemical compounds analogous to the vitamins and amino acids required for human nutrition, because plants are able to synthesize whatever compounds they need. They do require more than a dozen different chemical elements and these elements must be present in such forms as to allow an adequate availability for plant use. Within this restriction, nitrogen, for example, can be supplied with equal effectiveness in the form of urea, nitrates, ammonium compounds, or pure ammonia.

Virgin soil usually contains adequate amounts of all the elements required for proper plant nutrition. When a particular crop is grown on the same parcel of land year after year, however, the land may become exhausted of one or more specific nutrients. If such exhaustion occurs, nutrients in the form of fertilizers must be added to the soil. Plants can also be made to grow more lushly with suitable fertilizers.

Of the required nutrients, hydrogen, oxygen, and carbon are supplied in inexhaustible form by air and water. Sulfur, calcium, and iron are necessary nutrients that usually are present in soil in ample quantities. Lime (calcium) is often added to soil, but its function is primarily to reduce acidity and not, in the strict sense, to act as a fertilizer. Nitrogen is present in enormous quantities in the atmosphere, but plants are not able to use nitrogen in this form; bacteria provide nitrogen from the air to plants of the legume family through a process called nitrogen fixation. The three elements that most commonly must be supplied in fertilizers are nitrogen, phosphorus, and potassium. Certain other elements, such as boron, copper, and manganese, sometimes need to be included in small quantities.

Many fertilizers used since ancient times contain one or more of the three elements important to the soil. For example, manure and guano contain nitrogen. Bones contain small quantities of nitrogen and larger quantities of phosphorus. Wood ash contains appreciable quantities of potassium (depending considerably on the type of wood). Clover, alfalfa, and other legumes are grown as rotating crops and then plowed under, enriching the soil with nitrogen.

The term complete fertilizer often refers to any mixture containing all three important elements; such fertilizers are described by a set of three numbers. For example, 5-8-7 designates a fertilizer (usually in powder or granular form) containing 5 percent nitrogen, 8 percent phosphorus (calculated as phosphorus pentoxide), and 7 percent potassium (calculated as potassium oxide).

While fertilizers are essential to modern agriculture, their overuse can have harmful effects on plants and crops and on soil quality. In addition, the leaching of nutrients into bodies of water can lead to water pollution problems such as eutrophication, by causing excessive growth of vegetation.
The use of industrial waste materials in commercial fertilizers has been encouraged in the United States as a means of recycling waste products. The safety of this practice has recently been called into question. Its opponents argue that industrial wastes often contain elements that poison the soil and can introduce toxic chemicals into the food chain

37. PESTICIDES

Types of Pesticides:

A pesticide is any chemical which is used by man to control pests. The pests may be insects, plant diseases, fungi, weeds, nematodes, snails, slugs, etc. Therefore, insecticides, fungicides, herbicides, etc., are all types of pesticides. Some pesticides must only contact (touch) the pest to be deadly. Others must be swallowed to be effective. The way that each pesticide attacks a pest suggests the best way to apply it; to reach and expose all the pests. For example, a pesticide may be more effective and less costly as bait, rather than as a surface spray.

Insecticides:

Insecticides are chemicals used to control insects. Often the word "insecticide" is confused with the word "pesticide." It is, however, just one of many types of pesticides. An insecticide may kill the insect by touching it or it may have to be swallowed to be effective. Some insecticides kill both by touch and by swallowing. Insecticides called Systemic may be absorbed, injected, or fed into the plant or animal to be protected. When the insect feeds on this plant or animal, it ingests the systemic chemical and is killed.

Matricides and Acaroids:

Matricides (or Acaroids) are chemicals used to control mites (tiny Insecticides spider-like animals) and ticks. The chemicals usually must contact the mites or ticks to be effective. These animals are so numerous and small, that great care must be used to completely cover the area on which the mites live. Matricides are very similar in action to insecticides and often the same pesticide kills both insects and mites. The terms "broad spectrum," "short term," and "residual" are also used

Fungicides:

Fungicides are chemicals used to control the fungi which cause molds, rots, and plant diseases. All fungicides work by coming in contact with the fungus, because fungi do not "swallow" in the normal sense. Therefore, most fungicides are applied over a large surface area to try to directly hit every fungus. Some fungicides may be systemic in that the plant to be protected may be fed or injected with the chemical. The chemical then moves throughout the plant, killing the fungi. to describe matricides.

Herbicides:

Herbicides are chemicals used to control unwanted plants. These chemicals are a bit different from other pesticides because they are used to kill or slow the growth of some plants, rather than to protect them. Some herbicides kill every plant they contact, while others kill only certain plants.

Rodenticides:

Rodenticides are chemicals used to control rats, mice, bats and other rodents. Chemicals which control other mammals, birds, and fish are also grouped in this category by regulatory agencies. Most rodenticides are stomach poisons and are often applied as baits. Even rodenticides which act by contacting the pest are usually not applied over large surfaces because of the hazard to domestic animals or desirable wildlife. They are usually applied in limited areas such as runways, known feeding places, or as baits.

Nematicides:

Nematicides are chemicals used to control nematodes. Nematodes are tiny hair-like worms, many of which live in the soil and feed on plant roots. Very few of these worms live above ground. Usually, soil fumigants are used to control nematodes in the soil

38. MICROWAVE OVENS

The microwave oven is one of the great inventions of the 20th century .microwave ovens cook food in an amazingly short amount of time. A microwave oven uses microwaves to heat food. Microwaves are radio waves. In the case of microwave ovens, the commonly used radio wave frequency is roughly 2,500 megahertz (2.5 gigahertz). Radio waves in this frequency range have an interesting property: they are absorbed by water, fats and sugars. When they are absorbed they are converted directly into atomic motion -- heat. Microwaves in this frequency range have another interesting property: they are not absorbed by most plastics, glass or ceramics. Metal reflects microwaves, which is why metal pans do not work well in a microwave oven.

HAZARDS:

No doubt that microwave ovens have added many advantages to the daily lives but their frequent used has been negated by the doctors and physicians because of various serious health hazrads.Among the most serious of them is the cause of cancer.Infact some people have termed the microwaves as the "recipe of cancer."

Micro wave cooking is not natural and therefore it cannot be regarded as healthy. The wave radiations that are generated by a microwave oven during the process of cooking or heating any food item are considered extremely harmful and is conceived as one of the biggest resources of spreading the stomach and intestinal cancers.

The microwave exposure also reduces the nutritive value of the foods, loss of memory, emotional instability and a decrease of intelligence.

39. IMMUNIZATION

INTRODUCTION:

Immunization, also called vaccination or inoculation, a method of stimulating resistance in the human body to specific diseases using microorganisms—bacteria or viruses—that have been modified or killed. These treated microorganisms do not cause the disease, but rather trigger the body's immune system to build a defense mechanism that continuously guards against the disease. If a person immunized against a particular disease later comes into contact with the disease-causing agent, the immune system is immediately able to respond defensively.

Immunization has dramatically reduced the incidence of a number of deadly diseases. For example, a worldwide vaccination program resulted in the global eradication of smallpox in 1980, and in most developed countries immunization has essentially eliminated diphtheria, poliomyelitis, and neonatal tetanus. The number of cases of Homophiles influenza type b meningitis in the United States has dropped 95 percent among infants and children since 1988, when the vaccine for that disease was first introduced. In the United States, more than 90 percent of children receive all the recommended vaccinations by their second birthday. About 85 percent of Canadian children are immunized by age two.

TYPES OF IMMUNIZATION:

Scientists have developed two approaches to immunization: active immunization, which provides long-lasting immunity, and passive immunization, which gives temporary immunity. In active immunization, all or part of a disease-causing microorganism or a modified product of that microorganism is injected into the body to make the immune system respond defensively. Passive immunity is accomplished by injecting blood from an actively immunized human being or animal.

A -Active Immunization:

Vaccines that provide active immunization are made in a variety of ways, depending on the type of disease and the organism that causes it. The active components of the vaccinations are antigens, substances found in the disease-causing organism that the immune system recognizes as foreign. In response to the antigen, the immune system develops either antibodies or white blood cells called T lymphocytes, which are special attacker cells. Immunization mimics real infection but presents little or no risk to the recipient. Some immunizing agents provide complete protection against a disease for life. Other agents provide partial protection, meaning that the immunized person can contract the disease, but in a less severe form. These vaccines are usually considered risky for people who have a damaged immune system, such as those infected with the virus that causes acquired immunodeficiency syndrome (AIDS) or those receiving chemotherapy for cancer or organ transplantation. Without a healthy defense system to fight infection, these people may develop the disease that the vaccine is trying to prevent. Some immunizing agents require repeated inoculations—or booster shots—at specific intervals. Tetanus shots, for example, are recommended every ten years throughout life.

In order to make a vaccine that confers active immunization, scientists use an organism or part of one that has been modified so that it has a low risk of causing illness but still triggers the body's immune defenses against disease. One type of vaccine contains live organisms that have been attenuated—that is, their virulence has been weakened. This procedure is used to protect against yellow fever, measles, smallpox, and many other viral diseases.
Immunization can also occur when a person receives an injection of killed or inactivated organisms that are relatively harmless but that still contain antigens. This type of vaccination is used to protect against bacterial diseases such as poliomyelitis, typhoid fever, and diphtheria.

Some vaccines use only parts of an infectious organism that contain antigens, such as a protein cell wall or a flagellum. Known as cellular vaccines, they produce the desired immunity with a lower risk of producing potentially harmful immune reactions that may result from exposure to other parts of the organism. A cellular vaccines include the Haemophilus influenza type B vaccine for meningitis and newer versions of the whooping cough vaccine. Scientists use genetic engineering techniques to refine this approach further by isolating a gene or genes within an infectious organism that code for a particular antigen. The subunit vaccines produced by this method cannot cause disease and are safe to use in people who have an impaired immune system. Subunit vaccines for hepatitis B and pneumococcal infection, which causes pneumonia, became available in the late 1990s.

Active immunization can also be carried out using bacterial toxins that have been treated with chemicals so that they are no longer toxic, even though their antigens remain intact. This procedure uses the toxins produced by genetically engineered bacteria rather than the organism itself and is used in vaccinating against tetanus, botulism, and similar toxic diseases.

B -Passive Immunization:

Passive immunization is performed without injecting any antigen. In this method, vaccines contain antibodies obtained from the blood of an actively immunized human being or animal. The antibodies last for two to three weeks, and during that time the person is protected against the disease. Although short-lived, passive immunization provides immediate protection, unlike active immunization, which can take weeks to develop. Consequently, passive immunization can be lifesaving when a person has been infected with a deadly organism.

Occasionally there are complications associated with passive immunization. Diseases such as botulism and rabies once posed a particular problem. Immune globulin (antibody-containing plasma) for these diseases was once derived from the blood serum of horses. Although this animal material was specially treated before administration to humans, serious allergic reactions were common. Today, human-derived immune globulin is more widely available and the risk of side effects is reduced.

IMMUNIZATION RECOMMENDATIONS:

More than 50 vaccines for preventable diseases are licensed in the United States. The American Academy of Pediatrics and the U.S. Public Health Service recommend a series of immunizations beginning at birth. The initial series for children is complete by the time they reach the age of two, but booster vaccines are required for certain diseases, such as diphtheria and tetanus, in order to maintain adequate protection. When new vaccines are introduced, it is uncertain how long full protection will last. Recently, for example, it was discovered that a single injection of measles vaccine, first licensed in 1963 and administered to children at the age of 15 months, did not confer protection through adolescence and young adulthood. As a result, in the 1980s a series of measles epidemics occurred on college campuses throughout the United States among students who had been vaccinated as infants. To forestall future epidemics, health authorities now recommend that a booster dose of the measles, mumps, and rubella (also known as German measles) vaccine be administered at the time a child first enters school.
Not only children but also adults can benefit from immunization. Many adults in the United States are not sufficiently protected against tetanus, diphtheria, measles, mumps, and German measles. Health authorities recommend that most adults 65 years of age and older, and those with respiratory illnesses, be immunized against influenza (yearly) and pneumococcus (once).

HISTORY OF IMMUNIZATION:

The use of immunization to prevent disease predated the knowledge of both infection and immunology. In China in approximately 600 BC, smallpox material was inoculated through the nostrils. Inoculation of healthy people with a tiny amount of material from smallpox sores was first attempted in England in 1718and later in America. Those who survived the inoculation became immune to smallpox. American statesman Thomas Jefferson traveled from his home in Virginia to Philadelphia, Pennsylvania, to undergo this risky procedure.

A significant breakthrough came in 1796 when British physician Edward Jenner discovered that he could immunize patients against smallpox by inoculating them with material from cowpox sores. Cowpox is a far milder disease that, unlike smallpox, carries little risk of death or disfigurement. Jenner inserted matter from cowpox sores into cuts he made on the arm of a healthy eight-year-old boy. The boy caught cowpox. However, when Jenner exposed the boy to smallpox eight weeks later, the child did not contract the disease. The vaccination with cowpox had made him immune to the smallpox virus. Today we know that the cowpox virus antigens are so similar to those of the smallpox virus that they trigger the body's defenses against both diseases.

In 1885 Louis Pasteur created the first successful vaccine against rabies for a young boy who had been bitten 14 times by a rabid dog. Over the course of ten days, Pasteur injected progressively more virulent rabies organisms into the boy, causing the boy to develop immunity in time to avert death from this disease.

Another major milestone in the use of vaccination to prevent disease occurred with the efforts of two American physician-researchers. In 1954 Jonas Salk introduced an injectable vaccine containing an inactivated virus to counter the epidemic of poliomyelitis. Subsequently, Albert Sabin made great strides in the fight against this paralyzing disease by developing an oral vaccine containing a live weakened virus. Since the introduction of the polio vaccine, the disease has been nearly eliminated in many parts of the world.
As more vaccines are developed, a new generation of combined vaccines are becoming available that will allow physicians to administer a single shot for multiple diseases. Work is also under way to develop additional orally administered vaccines and vaccines for sexually transmitted diseases.

Possible future vaccines may include, for example, one that would temporarily prevent pregnancy. Such a vaccine would still operate by stimulating the immune system to recognize and attack antigens, but in this case the antigens would be those of the hormones that are necessary for pregnancy

40. FINGERPRINTING

INTRODUCTION:

Fingerprinting, method of identification using the impression made by the minute ridge formations or patterns found on the fingertips. No two persons have exactly the same arrangement of ridge patterns, and the patterns of any one individual remain unchanged through life. To obtain a set of fingerprints, the ends of the fingers are inked and then pressed or rolled one by one on some receiving surface. Fingerprints may be classified and filed on the basis of the ridge patterns, setting up an identification system that is almost infallible.

HISTORY:

The first recorded use of fingerprints was by the ancient Assyrians and Chinese for the signing of legal documents. Probably the first modern study of fingerprints was made by the Czech physiologist Johannes Evangelista Purkinje, who in 1823 proposed a system of classification that attracted little attention. The use of fingerprints for identification purposes was proposed late in the 19th century by the British scientist Sir Francis Galton, who wrote a detailed study of fingerprints in which he presented a new classification system using prints of all ten fingers, which is the basis of identification systems still in use. In the 1890s the police in Bengal, India, under the British police official Sir Edward Richard Henry, began using fingerprints to identify criminals. As assistant commissioner of metropolitan police, Henry established the first British fingerprint files in London in 1901. Subsequently, the use of fingerprinting as a means for identifying criminals spread rapidly throughout Europe and the United States, superseding the old Bertillon system of identification by means of body measurements.

MODERN USE:

As crime-detection methods improved, law enforcement officers found that any smooth, hard surface touched by a human hand would yield fingerprints made by the oily secretion present on the skin. When these so-called latent prints were dusted with powder or chemically treated, the identifying fingerprint pattern could be seen and photographed or otherwise preserved. Today, law enforcement agencies can also use computers to digitally record fingerprints and to transmit them electronically to other agencies for comparison. By comparing fingerprints at the scene of a crime with the fingerprint record of suspected persons, officials can establish absolute proof of the presence or identity of a person.

The confusion and inefficiency caused by the establishment of many separate fingerprint archives in the United States led the federal government to set up a central agency in 1924, the Identification Division of the Federal Bureau of Investigation (FBI). This division was absorbed in 1993 by the FBI's Criminal Justice Information Services Division, which now maintains the world's largest fingerprint collection. Currently the FBI has a library of more than 234 million civil and criminal fingerprint cards, representing 81 million people. In 1999 the FBI began full operation of the Integrated Automated Fingerprint Identification System (IAFIS), a computerized system that stores digital images of fingerprints for more than 36 million individuals, along with each individual's criminal history if one exists. Using IAFIS, authorities can conduct automated searches to identify people from their fingerprints and determine whether they have a criminal record. The system also gives state and local law enforcement agencies the ability to electronically transmit fingerprint information to the FBI. The implementation of IAFIS represented a breakthrough in crime fighting by reducing the time needed for fingerprint identification from weeks to minutes or hours.

41. INFRARED RADIATION

Infrared Radiation, emission of energy as electromagnetic waves in the portion of the spectrum just beyond the limit of the red portion of visible radiation (see Electromagnetic Radiation). The wavelengths of infrared radiation are shorter than those of radio waves and longer than those of light waves. They range between approximately 10-6 and 10-3 (about 0.0004 and 0.04 in).
Infrared radiation may be detected as heat, and instruments such as bolometer are used to detect it. See Radiation; Spectrum. Infrared radiation is used to obtain pictures of distant objects obscured by atmospheric haze, because visible light is scattered by haze but infrared radiation is not. The detection of infrared radiation is used by astronomers to observe stars and nebulas that are invisible in ordinary light or that emit radiation in the infrared portion of the spectrum.

An opaque filter that admits only infrared radiation is used for very precise infrared photographs, but an ordinary orange or lightred filter, which will absorb blue and violet light, is usually sufficient for most infrared pictures. Developed about 1880, infrared photography has today become an important diagnostic tool in medical science as well as in agriculture and industry. Use of infrared techniques reveals pathogenic conditions that are not visible to the eye or recorded on X-ray plates. Remote sensing by means of aerial and orbital infrared photography has been used to monitor crop conditions and insect and disease damage to large agricultural areas, and to locate mineral deposits. See Aerial Survey; Satellite, Artificial. In industry, infrared spectroscopy forms an increasingly important part of metal and alloy research, and infrared photography is used to monitor the quality of products. See also Photography: Photographic Films.

Infrared devices such as those used during World War II enable sharpshooters to see their targets in total visual darkness. These instruments consist essentially of an infrared lamp that sends out a beam of infrared radiation, often referred to as black light, and a telescope receiver that picks up returned radiation from the object and converts it to a visible image.

42. GREENHOUSE EFFECT

INTRODUCTION:

Greenhouse Effect, the capacity of certain gases in the atmosphere to trap heat emitted from the Earth's surface, thereby insulating and warming the Earth. Without the thermal blanketing of the natural greenhouse effect, the Earth's climate would be about 33 Celsius degrees (about 59 Fahrenheit degrees) cooler—too cold for most living organisms to survive.

The greenhouse effect has warmed the Earth for over 4 billion years. Now scientists are growing increasingly concerned that human activities may be modifying this natural process, with potentially dangerous consequences. Since the advent of the Industrial Revolution in the 1700s, humans have devised many inventions that burn fossil fuels such as coal, oil, and natural gas. Burning these fossil fuels, as well as other activities such as clearing land for agriculture or urban settlements, releases some of the same gases that trap heat in the atmosphere, including carbon dioxide, methane, and nitrous oxide. These atmospheric gases have risen to levels higher than at any time in the last 420,000 years. As these gases build up in the atmosphere, they trap more heat near the Earth's surface, causing Earth's climate to become warmer than it would naturally.

Scientists call this unnatural heating effect global warming and blame it for an increase in the Earth's surface temperature of about 0.6 Celsius degrees (about 1 Fahrenheit degree) over the last nearly 100 years. Without remedial measures, many scientists fear that global temperatures will rise 1.4 to 5.8 Celsius degrees (2.5 to 10.4 Fahrenheit degrees) by 2100. These warmer temperatures could melt parts of polar ice caps and most mountain glaciers, causing a rise in sea level of up to 1 m (40 in) within a century or two, which would flood coastal regions. Global warming could also affect weather patterns causing, among other problems, prolonged drought or increased flooding in some of the world's leading agricultural regions.

HOW THE GREENHOUSE EFFECT WORKS:

The greenhouse effect results from the interaction between sunlight and the layer of greenhouse gases in the Earth's atmosphere that extends up to 100 km (60 mi) above Earth's surface. Sunlight is composed of a range of radiant energies known as the solar spectrum, which includes visible light, infrared light, gamma rays, X rays, and ultraviolet light. When the Sun's radiation reaches the Earth's atmosphere, some 25 percent of the energy is reflected back into space by clouds and other atmospheric particles. About 20 percent is absorbed in the atmosphere. For instance, gas molecules in the uppermost layers of the atmosphere absorb the Sun's gamma rays and X rays. The Sun's ultraviolet radiation is absorbed by the ozone layer, located 19 to 48 km (12 to 30 mi) above the Earth's surface.

About 50 percent of the Sun's energy, largely in the form of visible light, passes through the atmosphere to reach the Earth's surface. Soils, plants, and oceans on the Earth's surface absorb about 85 percent of this heat energy, while the rest is reflected back into the atmosphere—most effectively by reflective surfaces such as snow, ice, and sandy deserts. In addition, some of the Sun's radiation that is absorbed by the Earth's surface becomes heat energy in the form of long-wave infrared radiation, and this energy is released back into the atmosphere.

Certain gases in the atmosphere, including water vapor, carbon dioxide, methane, and nitrous oxide, absorb this infrared radiant heat, temporarily preventing it from dispersing into space. As these atmospheric gases warm, they in turn emit infrared radiation in all directions. Some of this heat returns back to Earth to further warm the surface in what is known as the greenhouse effect, and some of this heat is eventually released to space. This heat transfer creates equilibrium between the total amount of heat that reaches the Earth from the Sun and the amount of heat that the Earth radiates out into space. This equilibrium or energy balance—the exchange of energy between the Earth's surface, atmosphere, and space—is important to maintain a climate that can support a wide variety of life.

The heat-trapping gases in the atmosphere behave like the glass of a greenhouse. They let much of the Sun's rays in, but keep most of that heat from directly escaping. Because of this, they are called greenhouse gases. Without these gases, heat energy absorbed and reflected from the Earth's surface would easily radiate back out to space, leaving the planet with an inhospitable temperature close to -19°C (2°F), instead of the present average surface temperature of 15°C (59°F).

To appreciate the importance of the greenhouse gases in creating a climate that helps sustain most forms of life, compare Earth to Mars and Venus. Mars has a thin atmosphere that contains low concentrations of heat-trapping gases. As a result, Mars has a weak greenhouse effect resulting in a largely frozen surface that shows no evidence of life. In contrast, Venus has an atmosphere containing high concentrations of carbon dioxide. This heat-trapping gas prevents heat radiated from the planet's surface from escaping into space, resulting in surface temperatures that average 462°C (864°F)—too hot to support life.

TYPES OF GREENHOUSE GASES:

Earth's atmosphere is primarily composed of nitrogen (78 percent) and oxygen (21 percent). These two most common atmospheric gases have chemical structures that restrict absorption of infrared energy. Only the few greenhouse gases, which make up less than 1 percent of the atmosphere, offer the Earth any insulation. Greenhouse gases occur naturally or are manufactured. The most abundant naturally occurring greenhouse gas is water vapor, followed by carbon dioxide, methane, and nitrous oxide. Human-made chemicals that act as greenhouse gases include chlorofluorocarbons (CFCs), hydro chlorofluorocarbons (HCFCs), and hydro fluorocarbons (HFCs).

Since the 1700s, human activities have substantially increased the levels of greenhouse gases in the atmosphere. Scientists are concerned that expected increases in the concentrations of greenhouse gases will powerfully enhance the atmosphere's capacity to retain infrared radiation, leading to an artificial warming of the Earth's surface.

A -Water Vapor

Water vapor is the most common greenhouse gas in the atmosphere, accounting for about 60 to 70 percent of the natural greenhouse effect. Humans do not have a significant direct impact on water vapor levels in the atmosphere. However, as human activities increase the concentration of other greenhouse gases in the atmosphere (producing warmer temperatures on Earth), the evaporation of oceans, lakes, and rivers, as well as water evaporation from plants, increase and raise the amount of water vapor in the atmosphere.

B -Carbon Dioxide

Carbon dioxide constantly circulates in the environment through a variety of natural processes known as the carbon cycle. Volcanic eruptions and the decay of plant and animal matter both release carbon dioxide into the atmosphere. In respiration, animals break down food to release the energy required to build and maintain cellular activity. A byproduct of respiration is the formation of carbon dioxide, which is exhaled from animals into the environment. Oceans, lakes, and rivers absorb carbon dioxide from the atmosphere. Through photosynthesis, plants collect carbon dioxide and use it to make their own food, in the process incorporating carbon into new plant tissue and releasing oxygen to the environment as a byproduct.

In order to provide energy to heat buildings, power automobiles, and fuel electricity-producing power plants, humans burn objects that contain carbon, such as the fossil fuels oil, coal, and natural gas; wood or wood products; and some solid wastes. When these products are burned, they release carbon dioxide into the air. In addition, humans cut down huge tracts of trees for lumber or to clear land for farming or building. This process, known as deforestation, can both release the carbon stored in trees and significantly reduce the number of trees available to absorb carbon dioxide.

As a result of these human activities, carbon dioxide in the atmosphere is accumulating faster than the Earth's natural processes can absorb the gas. By analyzing air bubbles trapped in glacier ice that is many centuries old, scientists have determined that carbon dioxide levels in the atmosphere have risen by 31 percent since 1750. And since carbon dioxide increases can remain in the atmosphere for centuries, scientists expect these concentrations to double or triple in the next century if current trends continue.

C -Methane

Many natural processes produce methane, also known as natural gas. Decomposition of carbon-containing substances found in oxygen-free environments, such as wastes in landfills, release methane. Ruminating animals such as cattle and sheep belch methane into the air as a byproduct of digestion. Microorganisms that live in damp soils, such as rice fields, produce methane when they break down organic matter. Methane is also emitted during coal mining and the production and transport of other fossil fuels.

Methane has more than doubled in the atmosphere since 1750, and could double again in the next century. Atmospheric concentrations of methane are far less than carbon dioxide, and methane only stays in the atmosphere for a decade or so. But scientists consider methane an extremely effective heat-trapping gas—one molecule of methane is 20 times more efficient at trapping infrared radiation radiated from the Earth's surface than a molecule of carbon dioxide.

D -Nitrous Oxide

Nitrous oxide is released by the burning of fossil fuels, and automobile exhaust is a large source of this gas. In addition, many farmers use nitrogen-containing fertilizers to provide nutrients to their crops. When these fertilizers break down in the soil, they emit nitrous oxide into the air. Plowing fields also releases nitrous oxide.

Since 1750 nitrous oxide has risen by 17 percent in the atmosphere. Although this increase is smaller than for the other greenhouse gases, nitrous oxide traps heat about 300 times more effectively than carbon dioxide and can stay in the atmosphere for a century.

E -Fluorinated Compounds

Some of the most potent greenhouse gases emitted are produced solely by human activities. Fluorinated compounds, including CFCs, HCFCs, and HFCs, are used in a variety of manufacturing processes. For each of these synthetic compounds, one molecule is several thousand times more effective in trapping heat than a single molecule of carbon dioxide.

CFCs, first synthesized in 1928, were widely used in the manufacture of aerosol sprays, blowing agents for foams and packing materials, as solvents, and as refrigerants. Nontoxic and safe to use in most applications, CFCs are harmless in the lower atmosphere. However, in the upper atmosphere, ultraviolet radiation breaks down CFCs, releasing chlorine into the atmosphere. In the mid-1970s, scientists began observing that higher concentrations of chlorine were destroying the ozone layer in the upper atmosphere. Ozone protects the Earth from harmful ultraviolet radiation, which can cause cancer and other damage to plants and animals. Beginning in 1987 with the Montréal Protocol on Substances that Deplete the Ozone Layer, representatives from 47 countries established control measures that limited the consumption of CFCs. By 1992 the Montréal Protocol was amended to completely ban the manufacture and use of CFCs worldwide, except in certain developing countries and for use in special medical processes such as asthma inhalers.

Scientists devised substitutes for CFCs, developing HCFCs and HFCs. Since HCFCs still release ozone-destroying chlorine in the atmosphere, production of this chemical will be phased out by the year 2030, providing scientists some time to develop a new generation of safer, effective chemicals. HFCs, which do not contain chlorine and only remain in the atmosphere for a short time, are now considered the most effective and safest substitute for CFCs.

F -Other Synthetic Chemicals

Experts are concerned about other industrial chemicals that may have heat-trapping abilities. In 2000 scientists observed rising concentrations of a previously unreported compound called trifluoromethyl sulphur pentafluoride. Although present in extremely low concentrations in the environment, the gas still poses a significant threat because it traps heat more effectively than all other known greenhouse gases. The exact sources of the gas, undisputedly produced from industrial processes, still remain uncertain.

OTHER FACTORS AFFECTING THE GREENHOUSE EFFECT:

Aerosols, also known as particulates, are airborne particles that absorb, scatter, and reflect radiation back into space. Clouds, windblown dust, and particles that can be traced to erupting volcanoes are examples of natural aerosols. Human activities, including the burning of fossil fuels and slash-and-burn farming techniques used to clear forestland, contribute additional aerosols to the atmosphere. Although aerosols are not considered a heat-trapping greenhouse gas, they do affect the transfer of heat energy radiated from the Earth to space. The effect of aerosols on climate change is still debated, but scientists believe that light-colored aerosols cool the Earth's surface, while dark aerosols like soot actually warm the atmosphere. The increase in global temperature in the last century is lower than many scientists predicted when only taking into account increasing levels of carbon dioxide, methane, nitrous oxide, and fluorinated compounds. Some scientists believe that aerosol cooling may be the cause of this unexpectedly reduced warming.

However, scientists do not expect that aerosols will ever play a significant role in offsetting global warming. As pollutants, aerosols typically pose a health threat, and the manufacturing or agricultural processes that produce them are subject to air-pollution control efforts. As a result, scientists do not expect aerosols to increase as fast as other greenhouse gases in the 21st century.

UNDERSTANDING THE GREENHOUSE EFFECT:

Although concern over the effect of increasing greenhouse gases is a relatively recent development, scientists have been investigating the greenhouse effect since the early 1800s. French mathematician and physicist Jean Baptist Joseph Fourier, while exploring how heat is conducted through different materials, was the first to compare the atmosphere to a glass vessel in 1827. Fourier recognized that the air around the planet lets in sunlight, much like a glass roof.

In the 1850s British physicist John Tyndall investigated the transmission of radiant heat through gases and vapors. Tyndall found that nitrogen and oxygen, the two most common gases in the atmosphere, had no heat-absorbing properties. He then went on to measure the absorption of infrared radiation by carbon dioxide and water vapor, publishing his findings in 1863 in a paper titled ―On Radiation through the Earth‗s Atmosphere.‖

Swedish chemist Svante August Arrhenius, best known for his Nobel Prize-winning work in electrochemistry, also advanced understanding of the greenhouse effect. In 1896 he calculated that doubling the natural concentrations of carbon dioxide in the atmosphere would increase global temperatures by 4 to 6 Celsius degrees (7 to 11 Fahrenheit degrees), a calculation that is not too far from today‗s estimates using more sophisticated methods. Arrhenius correctly predicted that when Earth‗s temperature warms, water vapor evaporation from the oceans increases. The higher concentration of water vapor in the atmosphere would then contribute to the greenhouse effect and global warming.

The predictions about carbon dioxide and its role in global warming set forth by Arrhenius were virtually ignored for over half a century, until scientists began to detect a disturbing change in atmospheric levels of carbon dioxide. In 1957 researchers at the Scripps Institution of Oceanography, based in San Diego, California, began monitoring carbon dioxide levels in the atmosphere from Hawaii‗s remote Mauna Loa Observatory located 3,000 m (11,000 ft) above sea level. When the study began, carbon dioxide concentrations in the Earth‗s atmosphere were 315 molecules of gas per million molecules of air (abbreviated parts per million or ppm). Each year carbon dioxide concentrations increased—to 323 ppm by 1970 and 335 ppm by 1980. By 1988 atmospheric carbon dioxide had increased to 350 ppm, an 8 percent increase in only 31 years.

As other researchers confirmed these findings, scientific interest in the accumulation of greenhouse gases and their effect on the environment slowly began to grow. In 1988 the World Meteorological Organization and the United Nations Environment Programmed established the Intergovernmental Panel on Climate Change (IPCC). The IPCC was the first international collaboration of scientists to assess the scientific, technical, and socioeconomic information related to the risk of human-induced climate change. The IPCC creates periodic assessment reports on advances in scientific understanding of the causes of climate change, its potential impacts, and strategies to control greenhouse gases. The IPCC played a critical role in establishing the United Nations Framework Convention on Climate Change (UNFCCC). The UNFCCC, which provides an international policy framework for addressing climate change issues, was adopted by the United Nations General Assembly in 1992.

Today scientists around the world monitor atmospheric greenhouse gas concentrations and create forecasts about their effects on global temperatures. Air samples from sites spread across the globe are analyzed in laboratories to determine levels of individual greenhouse gases. Sources of greenhouse gases, such as automobiles, factories, and power plants, are monitored directly to determine their emissions. Scientists gather information about climate systems and use this information to create and test computer models that simulate how climate could change in response to changing conditions on the Earth and in the atmosphere. These models act as high-tech crystal balls to project what may happen in the future as greenhouse gas levels rise. Models can only provide approximations, and some of the predictions based on these models often spark controversy within the science community. Nevertheless, the basic concept of global warming is widely accepted by most climate scientists.

EFFORTS TO CONTROL GREENHOUSE GASES:

Due to overwhelming scientific evidence and growing political interest, global warming is currently recognized as an important national and international issue. Since 1992 representatives from over 160 countries have met regularly to discuss how to reduce worldwide greenhouse gas emissions. In 1997 representatives met in Kyôto, Japan, and produced an agreement, known as the Kyôto Protocol, which requires industrialized countries to reduce their emissions by 2012 to an average of 5 percent below 1990 levels. To help countries meet this agreement cost-effectively, negotiators are trying to develop a system in which nations that have no obligations or that have successfully met their reduced emissions obligations could profit by selling or trading their extra emissions quotas to other countries that are struggling to reduce their emissions. Negotiating such detailed emissions trading rules has been a contentious task for the world community since the signing of the Kyôto Protocol. A ratified agreement is still not yet in force, and ratification received a setback in 2001 when newly elected U.S. president George W. Bush renounced the treaty on the grounds that the required carbon-dioxide reductions in the United States would be too costly. He also objected that developing nations would not be bound by similar carbon-dioxide reducing obligations. However, many experts expect that as the scientific evidence about the dangers of global warming continues to mount, nations will be motivated to cooperate more effectively to reduce the risks of climate change.

43. ANTIMATTER

Antimatter, matter composed of elementary particles that are, in a special sense, mirror images of the particles that make up ordinary matter as it is known on earth. Antiparticles have the same mass as their corresponding particles but have opposite electric charges or other properties related to electromagnetism. For example, the antimatter electron, or positron, has opposite electric charge and magnetic moment (a property that determines how it behaves in a magnetic field), but is identical in all other respects to the electron. The antimatter equivalent of the charge less neutron, on the other hand, differs in having a magnetic moment of opposite sign (magnetic moment is another electromagnetic property). In all of the other parameters involved in the dynamical properties of elementary particles, such as mass, spin, and partial decay, antiparticles are identical with their corresponding particles.

The existence of antiparticles was first proposed by the British physicist Paul Adrian Maurice Dirac, arising from his attempt to apply the techniques of relativistic mechanics (see Relativity) to quantum theory. In 1928 he developed the concept of a positively charged electron but its actual existence was established experimentally in 1932. The existence of other antiparticles was presumed but not confirmed until 1955, when antiprotons and antineutrons were observed in particle accelerators. Since then, the full range of antiparticles has been observed or indicated. Antimatter atoms were created for the first time in September 1995 at the European Organization for Nuclear Research (CERN). Positrons were combined with antimatter protons to produce antimatter hydrogen atoms. These atoms of antimatter exist only for forty-billionths of a second, but physicists hope future experiments will determine what differences there are between normal hydrogen and its antimatter counterpart.

A profound problem for particle physics and for cosmology in general is the apparent scarcity of antiparticles in the universe. Their nonexistence, except momentarily, on earth is understandable, because particles and antiparticles are mutually annihilated with a great release of energy when they meet (see Annihilation). Distant galaxies could possibly be made of antimatter, but no direct method of confirmation exists. Most of what is known about the far universe arrives in the form of photons, which are identical with their antiparticles and thus reveal little about the nature of their sources. The prevailing opinion, however, is that the universe consists overwhelmingly of ―ordinary‖ matter, and explanations for this have been proposed by recent cosmological theory (see Inflationary Theory).

In 1997 scientists studying data gathered by the Compton Gamma Ray Observatory (GRO) operated by the National Aeronautics and Space Administration (NASA) found that the earth's home galaxy—the Milky Way—contains large clouds of antimatter particles. Astronomers suggest that these clouds form when high-energy events—such as the collision of neutron stars, exploding stars, or black holes—create radioactive elements that decay into matter and antimatter or heat matter enough to make it split into particles of matter and antimatter. When antimatter particles meet particles of matter, the two annihilate each other and produce a burst of gamma rays. It was these gamma rays that GRO detected

44. MAGMA

INTRODUCTION:

Magma, molten or partially molten rock beneath the earth's surface. Magma is generated when rock deep underground melts due to the high temperatures and pressures inside the earth. Because magma is lighter than the surrounding rock, it tends to rise. As it moves upward, the magma encounters colder rock and begins to cool. If the temperature of the magma drops low enough, the magma will crystallize underground to form rock; rock that forms in this way is called intrusive, or plutonic igneous rock, as the magma has formed by intruding the surrounding rocks. If the crust through which the magma passes is sufficiently shallow, warm, or fractured, and if the magma is sufficiently hot and fluid, the magma will erupt at the surface of the earth, possibly forming volcanoes. Magma that erupts is called lava.

COMPOSITION OF MAGMA:

Magmas are liquids that contain a variety of melted minerals and dissolved gases. Because magmas form deep underground, however, geologists cannot directly observe and measure their original composition. This difficulty has led to controversy over the exact chemical composition of magmas. Geologists cannot simply assume it is the same as the composition of the rock in the source region. One reason for this is that the source rock may melt only partially, releasing only the minerals with the lowest melting points. For this reason, the composition of magma produced by melting 1 percent of a rock is different from the composition of magma produced by melting 20 percent of a rock. Experiments have shown that the temperature and pressure of the location within the earth, and the amount of water present at that location affect the amount of melting. Because temperature and pressure increase as depth within the earth increases, melting an identical source rock at different depths will produce magmas of different composition. Combining these considerations with the fact that the composition of the source rock may be different in different geographic regions, there is a considerable range of possible compositions for magma.

As magma moves toward the surface, the pressure and temperature decrease, which causes partial crystallization, or the formation of mineral crystals within the magma. The compositions of the minerals that crystallize are different from the initial composition of the magma because of changes in temperature and pressure, hence the composition of the remaining liquid changes. The resultant crystals may separate from the liquid either by sinking or by a process known as filter-pressing, in which pressure compresses the liquid and causes it to move toward regions of lower pressure while leaving the crystals behind. As a result, the composition of the remaining magma is different from that of the initial magma. This process is known as magmatic differentiation, and is the principal mechanism whereby a wide variety of magmas and rocks can be produced from a single primary magma (see Igneous Rock: Formation of Igneous Rocks).

The composition of magma can also be modified by chemical interactions with, and melting of, the rocks through which it passes on its way upward. This process is known as assimilation. Magma cannot usually supply enough heat to melt a large amount of the surrounding rock, so assimilation seldom produces a significant change in the composition of magma.

Magmas also contain dissolved gases, because gases are especially soluble (easily dissolved) in liquids when the liquids are under pressure. Magma deep underground is under thousands of atmospheres (units of measure) of pressure due to the weight of the overlying rock. Gases commonly dissolved in magma are carbon dioxide, water vapor, and sulfur dioxide.

PHYSICAL PROPERTIES OF MAGMA:

The density and viscosity, or thickness, of magma is key physical factors that affect its upward passage. Most rocks expand about 10 percent when they melt, and hence most magma has a density of about 90 percent of the equivalent solid rock. This density difference produces sufficient buoyancy in the magma to cause it to rise toward the surface.

The viscosity of a fluid is a measure of its resistance to flow. The viscosity of a magma affects how quickly the magma will rise, and it determines whether crystals of significantly different density will sink rapidly enough to change the bulk composition of the magma. Viscosity also influences the rate of release of gases from the magma when pressure is released. The viscosity of magma is closely related to the magma's chemical composition. Magma rich in silicon and poor in magnesium and iron, called felsic magma, is very viscous, or thick (see Igneous Rock: Felsic Rocks). Magma poor in silicon and rich in magnesium and iron, called mafic magma, is quite fluid (see Igneous Rock: Mafic Rocks).

GEOLOGICAL FEATURES FORMED BY MAGMA:

Some magma reaches the surface of the earth and erupts from volcanoes or fissures before they solidify. Other magmas fail to reach the surface before they solidify. Magma that reaches the surface and is erupted, or extruded, forms extrusive igneous rocks. Magma that intrudes, or pushes its way into rocks deep underground and solidifies there forms intrusive igneous rock. Volcanoes are cone-shaped mountains formed by the eruption of lava. Magma collects in a reservoir surrounded by rock, called a magma chamber, about 10 to 20 km (6 to 12 mi) below the volcano. A conduit known as a volcanic pipe provides a passage for the magma from the magma chamber to the volcano. As the magma rises in the conduit, the pressure of the overlying rock drops. Gases expand and bubble out that was kept dissolved in the magma by the pressure. The rapidly expanding gases propel the magma up the volcanic pipe, forcing the magma to the surface and leading to an eruption. The same process occurs when a shaken bottle of soda is suddenly opened.

The viscosity and dissolved-gas content of the magma control the character of the eruption. Low-viscosity magmas often have low gas content. They flow easily from volcanic conduits and result in relatively quiet eruptions. Once the magma reaches the surface, it rapidly spreads out and over the volcano. Such fluid lava creates broad, gently sloped volcanoes called shield volcanoes, so called because they resemble giant shields lying on the ground.
Low-viscosity lava can also flow from fissures (long cracks in the rock), forming huge lava lakes. Repeated eruptions result in formations called flood basalts. The Columbia Plateau, in the states of Washington, Oregon, and Idaho, is a flood basalt that covers nearly 200,000 sq km (about 80,000 sq mi) and is more than 4000 m (13,000 ft) thick in places.

If low-viscosity magma contains moderate amounts of dissolved gas, the released gases can eject the magma from the top of the volcano with enough force to form a lava fountain. The blobs of lava that are ejected into the air are called pyroclasts. They accumulate around the base of the fountain, forming a cinder cone.

Medium-viscosity magmas usually contain higher amounts of gases. They tend to form stratovolcanoes. The higher amounts of gases in the magma lead to very explosive eruptions that spew out large amounts of volcanic material. Stratovolcanoes have steeper sides than shield volcanoes. They are also known as composite volcanoes because they are made up of alternating layers of lava flows and deposits of pyroclasts.

High-viscosity magmas do not extrude easily though volcanic conduits. They often have a high gas content that can cause catastrophic eruptions. Both of these properties tend to promote explosive behavior, such as occurred on May 18, 1980 at Mount Saint Helens in Washington, when about 400 m (about 1300 ft) of rock was blasted off of its summit.

Intrusive bodies of rock formed from magma are classified by their size and shape. Batholith is an intrusive body that covers more than 100 sq km (nearly 40 sq mi). Lopoliths are saucer-shaped intrusions and may be up to 100 km (60 mi) in diameter and 8 km (5 mi) thick. Laccoliths have a flat base and a domed ceiling and are usually smaller than lopoliths. Sills and dikes are sheetlike intrusions that are very thin relative to their length. They can be less than one meter (about one yard) to several hundred meters thick but can be larger; the Palisades sill in the state of New York is 300 m (1000 ft) thick and 80 km (50 mi) long. Sills are formed when magma is forced between beds of layered rock; they run parallel to the layering of the surrounding rock. Dikes are formed when magma is forced into cracks in the surrounding rock; they tend to run perpendicular to the layering of the surrounding rock.

45. Brain

I -INTRODUCTION:

Brain, portion of the central nervous system contained within the skull. The brain is the control center for movement, sleep, hunger, thirst, and virtually every other vital activity necessary to survival. All human emotions—including love, hate, fear, anger, elation, and sadness—are controlled by the brain. It also receives and interprets the countless signals that are sent to it from other parts of the body and from the external environment. The brain makes us conscious, emotional, and intelligent.

II -ANATOMY:

The adult human brain is a 1.3-kg (3-lb) mass of pinkish-gray jellylike tissue made up of approximately 100 billion nerve cells, or neurons; neuralgia (supporting-tissue) cells; and vascular (blood-carrying) and other tissues.
Between the brain and the cranium—the part of the skull that directly covers the brain—are three protective membranes, or meninges. The outermost membrane, the dura mater, is the toughest and thickest. Below the dura mater is a middle membrane,

called the arachnoids layer. The innermost membrane, the pie mater, consists mainly of small blood vessels and follows the contours of the surface of the brain.

A clear liquid, the cerebrospinal fluid, bathes the entire brain and fills a series of four cavities, called ventricles, near the center of the brain. The cerebrospinal fluid protects the internal portion of the brain from varying pressures and transports chemical substances within the nervous system.

From the outside, the brain appears as three distinct but connected parts: the cerebrum (the Latin word for brain)—two large, almost symmetrical hemispheres; the cerebellum (―little brain‖)—two smaller hemispheres located at the back of the cerebrum; and the brain stem—a central core that gradually becomes the spinal cord, exiting the skull through an opening at its base called the foramen magnum. Two other major parts of the brain, the thalamus and the hypothalamus, lie in the midline above the brain stem underneath the cerebellum.

The brain and the spinal cord together make up the central nervous system, which communicates with the rest of the body through the peripheral nervous system. The peripheral nervous system consists of 12 pairs of cranial nerves extending from the cerebrum and brain stem; a system of other nerves branching throughout the body from the spinal cord; and the autonomic nervous system, which regulates vital functions not under conscious control, such as the activity of the heart muscle, smooth muscle (involuntary muscle found in the skin, blood vessels, and internal organs), and glands.

A -Cerebrum

Most high-level brain functions take place in the cerebrum. Its two large hemispheres make up approximately 85 percent of the brain's weight. The exterior surface of the cerebrum, the cerebral cortex, is a convoluted, or folded, grayish layer of cell bodies known as the gray matter. The gray matter covers an underlying mass of fibers called the white matter. The convolutions are made up of ridgelike bulges, known as gyri, separated by small grooves called sulci and larger grooves called fissures. Approximately two-thirds of the cortical surface is hidden in the folds of the sulci. The extensive convolutions enable a very large surface area of brain cortex—about 1.5 m2 (16 ft2) in an adult—to fit within the cranium. The pattern of these convolutions is similar, although not identical, in all humans.

The two cerebral hemispheres are partially separated from each other by a deep fold known as the longitudinal fissure. Communication between the two hemispheres is through several concentrated bundles of axons, called commissures, the largest of which is the corpus callosum.

Several major sulci divide the cortex into distinguishable regions. The central sulcus, or Rolandic fissure, runs from the middle of the top of each hemisphere downward, forward, and toward another major sulcus, the lateral (―side‖), or Sylvian, sulcus. These and other sulci and gyri divide the cerebrum into five lobes: the frontal, parietal, temporal, and occipital lobes and the insula.

The frontal lobe is the largest of the five and consists of all the cortex in front of the central sulcus. Broca's area, a part of the cortex related to speech, is located in the frontal lobe. The parietal lobe consists of the cortex behind the central sulcus to a sulcus near the back of the cerebrum known as the parieto-occipital sulcus. The parieto-occipital sulcus, in turn, forms the front border of the occipital lobe, which is the rearmost part of the cerebrum. The temporal lobe is to the side of and below the lateral sulcus. Wernicke's area, a part of the cortex related to the understanding of language, is located in the temporal lobe. The insula lies deep within the folds of the lateral sulcus.

The cerebrum receives information from all the sense organs and sends motor commands (signals that result in activity in the muscles or glands) to other parts of the brain and the rest of the body. Motor commands are transmitted by the motor cortex, a strip of cerebral cortex extending from side to side across the top of the cerebrum just in front of the central sulcus. The sensory cortex, a parallel strip of cerebral cortex just in back of the central sulcus, receives input from the sense organs.

Many other areas of the cerebral cortex have also been mapped according to their specific functions, such as vision, hearing, speech, emotions, language, and other aspects of perceiving, thinking, and remembering. Cortical regions known as associative cortex are responsible for integrating multiple inputs, processing the information, and carrying out complex responses.

B -rebellum

The cerebellum coordinates body movements. Located at the lower back of the brain beneath the occipital lobes, the cerebellum is divided into two lateral (side-by-side) lobes connected by a fingerlike bundle of white fibers called the vermis. The outer layer, or cortex, of the cerebellum consists of fine folds called folia. As in the cerebrum, the outer layer of cortical gray matter surrounds a deeper layer of white matter and nuclei (groups of nerve cells). Three fiber bundles called cerebellar peduncles connect the cerebellum to the three parts of the brain stem—the midbrain, the pons, and the medulla oblongata.

The cerebellum coordinates voluntary movements by fine-tuning commands from the motor cortex in the cerebrum. The cerebellum also maintains posture and balance by controlling muscle tone and sensing the position of the limbs. All motor activity, from hitting a baseball to fingering a violin, depends on the cerebellum.

C -Thalamus and Hypothalamus

the thalamus and the hypothalamus lie underneath the cerebrum and connect it to the brain stem. The thalamus consists of two rounded masses of gray tissue lying within the middle of the brain, between the two cerebral hemispheres. The thalamus is the main relay station for incoming sensory signals to the cerebral cortex and for outgoing motor signals from it. All sensory input to the brain, except that of the sense of smell, connects to individual nuclei of the thalamus.

The hypothalamus lies beneath the thalamus on the midline at the base of the brain. It regulates or is involved directly in the control of many of the body's vital drives and activities, such as eating, drinking, temperature regulation, sleep, emotional behavior, and sexual activity. It also controls the function of internal body organs by means of the autonomic nervous system, interacts closely with the pituitary gland, and helps coordinate activities of the brain stem.

D -Brain Stem
The brain stem is evolutionarily the most primitive part of the brain and is responsible for sustaining the basic functions of life, such as breathing and blood pressure. It includes three main structures lying between and below the two cerebral hemispheres— the midbrain, pons, and medulla oblongata.

D1 -Midbrain
The topmost structure of the brain stem is the midbrain. It contains major relay stations for neurons transmitting signals to the cerebral cortex, as well as many reflex centers—pathways carrying sensory (input) information and motor (output) commands. Relay and reflex centers for visual and auditory (hearing) functions are located in the top portion of the midbrain. A pair of nuclei called the superior colliculus control reflex actions of the eye, such as blinking, opening and closing the pupil, and focusing the lens. A second pair of nuclei, called the inferior colliculus, control auditory reflexes, such as adjusting the ear to the volume of sound. At the bottom of the midbrain are reflex and relay centers relating to pain, temperature, and touch, as well as several regions associated with the control of movement, such as the red nucleus and the substantial nigra.

D2 -Pons
Continuous with and below the midbrain and directly in front of the cerebellum is a prominent bulge in the brain stem called the pons. The pons consists of large bundles of nerve fibers that connect the two halves of the cerebellum and also connect each side of the cerebellum with the opposite-side cerebral hemisphere. The pons serves mainly as a relay station linking the cerebral cortex and the medulla oblongata.

D3 -Medulla Oblongata
The long, stalklike lowermost portion of the brain stem is called the medulla oblongata. At the top, it is continuous with the pons and the midbrain; at the bottom, it makes a gradual transition into the spinal cord at the foramen magnum. Sensory and motor nerve fibers connecting the brain and the rest of the body cross over to the opposite side as they pass through the medulla. Thus, the left half of the brain communicates with the right half of the body and the right half of the brain with the left half of the body.

D4 -Reticular Formation
running up the brain stem from the medulla oblongata through the pons and the midbrain is a netlike formation of nuclei known as the reticular formation. The reticular formation controls respiration, cardiovascular function (see Heart), digestion, levels of alertness, and patterns of sleep. It also determines which parts of the constant flow of sensory information into the body are received by the cerebrum.

E -Brain Cells
There are two main types of brain cells: neurons and neuralgia. Neurons are responsible for the transmission and analysis of all electrochemical communication within the brain and other parts of the nervous system. Each neuron is composed of a cell body called a soma, a major fiber called an axon, and a system of branches called dendrites. Axons, also called nerve fibers, convey electrical signals away from the soma and can be up to 1 m (3.3 ft) in length. Most axons are covered with a protective sheath of myelin, a substance made of fats and protein, which insulates the axon. Myelinated axons conduct neuronal signals faster than do unmyelinated axons. Dendrites convey electrical signals toward the soma, are shorter than axons, and are usually multiple and branching.

Neuroglial cells are twice as numerous as neurons and account for half of the brain's weight. Neuroglia (from glia, Greek for —glue‖) provides structural support to the neurons. Neuroglial cells also form myelin, guide developing neurons, take up chemicals involved in cell-to-cell communication, and contribute to the maintenance of the environment around neurons.

F -Cranial Nerves
Twelve pairs of cranial nerves arise symmetrically from the base of the brain and are numbered, from front to back, in the order in which they arise. They connect mainly with structures of the head and neck, such as the eyes, ears, nose, mouth, tongue, and throat. Some are motor nerves, controlling muscle movement; some are sensory nerves, conveying information from the sense organs; and others contain fibers for both sensory and motor impulses. The first and second pairs of cranial nerves—the olfactory (smell) nerve and the optic (vision) nerve—carry sensory information from the nose and eyes, respectively, to the undersurface of the cerebral hemispheres. The other ten pairs of cranial nerves originate in or end in the brain stem.

III -HOW THE BRAIN WORKS:

the brain functions by complex neuronal, or nerve cell, circuits (see Neurophysiology). Communication between neurons is both electrical and chemical and always travels from the dendrites of a neuron, through its soma, and out its axon to the dendrites of another neuron.
Dendrites of one neuron receive signals from the axons of other neurons through chemicals known as neurotransmitters. The neurotransmitters set off electrical charges in the dendrites, which then carry the signals electrochemically to the soma. The soma integrates the information, which is then transmitted electrochemically down the axon to its tip.

At the tip of the axon, small, bubblelike structures called vesicles release neurotransmitters that carry the signal across the synapse, or gap, between two neurons. There are many types of neurotransmitters, including norepinephrine, dopamine, and serotonin. Neurotransmitters can be excitatory (that is, they excite an electrochemical response in the dendrite receptors) or inhibitory (they block the response of the dendrite receptors).
One neuron may communicate with thousands of other neurons, and many thousands of neurons are involved with even the simplest behavior. It is believed that these connections and their efficiency can be modified, or altered, by experience.

Scientists have used two primary approaches to studying how the brain works. One approach is to study brain function after parts of the brain have been damaged. Functions that disappear or that are no longer normal after injury to specific regions of the brain can often be associated with the damaged areas. The second approach is to study the response of the brain to direct stimulation or to stimulation of various sense organs.

Neurons are grouped by function into collections of cells called nuclei. These nuclei are connected to form sensory, motor, and other systems. Scientists can study the function of somatosensory (pain and touch), motor, olfactory, visual, auditory, language, and other systems by measuring the physiological (physical and chemical) changes that occur in the brain when these senses are activated. For example, electroencephalography (EEG) measures the electrical activity of specific groups of neurons through electrodes attached to the surface of the skull. Electrodes inserted directly into the brain can give readings of individual neurons. Changes in blood flow, glucose (sugar), or oxygen consumption in groups of active cells can also be mapped.

Although the brain appears symmetrical, how it functions is not. Each hemisphere is specialized and dominates the other in certain functions. Research has shown that hemispheric dominance is related to whether a person is predominantly right-handed or left-handed (see Handedness). In most right-handed people, the left hemisphere processes arithmetic, language, and speech. The right hemisphere interprets music, complex imagery, and spatial relationships and recognizes and expresses emotion. In lefthanded people, the pattern of brain organization is more variable.

Hemispheric specialization has traditionally been studied in people who have sustained damage to the connections between the two hemispheres, as may occur with stroke, an interruption of blood flow to an area of the brain that causes the death of nerve cells in that area. The division of functions between the two hemispheres has also been studied in people who have had to have the connection between the two hemispheres surgically cut in order to control severe epilepsy, a neurological disease characterized by convulsions and loss of consciousness.

A -Vision
The visual system of humans is one of the most advanced sensory systems in the body (see Vision). More information is conveyed visually than by any other means. In addition to the structures of the eye itself, several cortical regions—collectively called primary visual and visual associative cortex—as well as the midbrain is involved in the visual system. Conscious processing of visual input occurs in the primary visual cortex, but reflexive—that is, immediate and unconscious—responses occur at the superior colliculus in the midbrain. Associative cortical regions—specialized regions that can associate, or integrate, multiple inputs—in the parietal and frontal lobes along with parts of the temporal lobe are also involved in the processing of visual information and the establishment of visual memories.

B -Language
Language involves specialized cortical regions in a complex interaction that allows the brain to comprehend and communicate abstract ideas. The motor cortex initiates impulses that travel through the brain stem to produce audible sounds. Neighboring regions of motor cortex, called the supplemental motor cortex, are involved in sequencing and coordinating sounds. Broca's area of the frontal lobe is responsible for the sequencing of language elements for output. The comprehension of language is dependent upon Wernicke's area of the temporal lobe. Other cortical circuits connect these areas.

C -Memory
Memory is usually considered a diffusely stored associative process—that is, it puts together information from many different sources. Although research has failed to identify specific sites in the brain as locations of individual memories, certain brain areas are critical for memory to function. Immediate recall—the ability to repeat short series of words or numbers immediately after hearing them—is thought to be located in the auditory associative cortex. Short-term memory—the ability to retain a limited amount of information for up to an hour—is located in the deep temporal lobe. Long-term memory probably involves exchanges between the medial temporal lobe, various cortical regions, and the midbrain.

D -The Autonomic Nervous System
The autonomic nervous system regulates the life support systems of the body reflexively—that is, without conscious direction. It automatically controls the muscles of the heart, digestive system, and lungs; certain glands; and homeostasis—that is, the equilibrium of the internal environment of the body (see Physiology). The autonomic nervous system itself is controlled by nerve centers in the spinal cord and brain stem and is fine-tuned by regions higher in the brain, such as the midbrain and cortex. Reactions such as blushing indicate that cognitive, or thinking, centers of the brain are also involved in autonomic responses.

IV -BRAIN DISORDERS:

the brain is guarded by several highly developed protective mechanisms. The bony cranium, the surrounding meninges, and the cerebrospinal fluid all contribute to the mechanical protection of the brain. In addition, a filtration system called the blood-brain barrier protects the brain from exposure to potentially harmful substances carried in the bloodstream.
Brain disorders have a wide range of causes, including head injury, stroke, bacterial diseases, complex chemical imbalances, and changes associated with aging.

A -Head Injury
Head injury can initiate a cascade of damaging events. After a blow to the head, a person may be stunned or may become unconscious for a moment.
This injury, called a concussion, usually leaves no permanent damage. If the blow is more severe and hemorrhage (excessive bleeding) and swelling occur, however, severe headache, dizziness, paralysis, a convulsion, or temporary blindness may result, depending on the area of the brain affected. Damage to the cerebrum can also result in profound personality changes.
Damage to Broca's area in the frontal lobe causes difficulty in speaking and writing, a problem known as Broca's aphasia. Injury to Wernicke's area in the left temporal lobe results in an inability to comprehend spoken language, called Wernicke's aphasia.

An injury or disturbance to a part of the hypothalamus may cause a variety of different symptoms, such as loss of appetite with an extreme drop in body weight; increase in appetite leading to obesity; extraordinary thirst with excessive urination (diabetes insipidus); failure in body-temperature control, resulting in either low temperature (hypothermia) or high temperature (fever); excessive emotionality; and uncontrolled anger or aggression. If the relationship between the hypothalamus and the pituitary gland is damaged (see Endocrine System), other vital bodily functions may be disturbed, such as sexual function, metabolism, and cardiovascular activity.
Injury to the brain stem is even more serious because it houses the nerve centers that control breathing and heart action. Damage to the medulla oblongata usually results in immediate death.

B -Stroke
A stroke is damage to the brain due to an interruption in blood flow. The interruption may be caused by a blood clot (see Embolism; Thrombosis), constriction of a blood vessel, or rupture of a vessel accompanied by bleeding. A pouchlike expansion of the wall of a blood vessel, called an aneurysm, may weaken and burst, for example, because of high blood pressure.

Sufficient quantities of glucose and oxygen, transported through the bloodstream, are needed to keep nerve cells alive. When the blood supply to a small part of the brain is interrupted, the cells in that area die and the function of the area is lost. A massive stroke can cause a one-sided paralysis (hemiplegia) and sensory loss on the side of the body opposite the hemisphere damaged by the stroke.

C -Brain Diseases
Epilepsy is a broad term for a variety of brain disorders characterized by seizures, or convulsions. Epilepsy can result from a direct injury to the brain at birth or from a metabolic disturbance in the brain at any time later in life.
Some brain diseases, such as multiple sclerosis and Parkinson disease, are progressive, becoming worse over time. Multiple sclerosis damages the myelin sheath around axons in the brain and spinal cord. As a result, the affected axons cannot transmit nerve impulses properly. Parkinson disease destroys the cells of the substantial nigra in the midbrain, resulting in a deficiency in the neurotransmitter dopamine that affects motor functions.

Cerebral palsy is a broad term for brain damage sustained close to birth that permanently affects motor function. The damage may take place either in the developing fetus, during birth, or just after birth and is the result of the faulty development or breaking down of motor pathways. Cerebral palsy is non progressive—that is, it does not worsen with time.
A bacterial infection in the cerebrum (see Encephalitis) or in the coverings of the brain (see Meningitis), swelling of the brain (see Edema), or an abnormal growth of healthy brain tissue (see Tumor) can all cause an increase in intracranial pressure and result in serious damage to the brain.

Scientists are finding that certain brain chemical imbalances are associated with mental disorders such as schizophrenia and depression. Such findings have changed scientific understanding of mental health and have resulted in new treatments that chemically correct these imbalances.
During childhood development, the brain is particularly susceptible to damage because of the rapid growth and reorganization of nerve connections. Problems that originate in the immature brain can appear as epilepsy or other brain-function problems in adulthood.

Several neurological problems are common in aging. Alzheimer's disease damages many areas of the brain, including the frontal, temporal, and parietal lobes. The brain tissue of people with Alzheimer's disease shows characteristic patterns of damaged neurons, known as plaques and tangles. Alzheimer's disease produces a progressive dementia (see Senile Dementia), characterized by symptoms such as failing attention and memory, loss of mathematical ability, irritability, and poor orientation in space and time.

V -BRAIN IMAGING:

Several commonly used diagnostic methods give images of the brain without invading the skull. Some portray anatomy—that is, the structure of the brain—whereas others measure brain function. Two or more methods may be used to complement each other, together providing a more complete picture than would be possible by one method alone.
Magnetic resonance imaging (MRI), introduced in the early 1980s, beams high-frequency radio waves into the brain in a highly magnetized field that causes the protons that form the nuclei of hydrogen atoms in the brain to reemit the radio waves. The reemitted radio waves are analyzed by computer to create thin cross-sectional images of the brain. MRI provides the most detailed images of the brain and is safer than imaging methods that use X rays. However, MRI is a lengthy process and also cannot be used with people who have pacemakers or metal implants, both of which are adversely affected by the magnetic field.

Computed tomography (CT), also known as CT scans, developed in the early 1970s. This imaging method X-rays the brain from many different angles, feeding the information into a computer that produces a series of cross-sectional images. CT is particularly useful for diagnosing blood clots and brain tumors. It is a much quicker process than magnetic resonance imaging and is therefore advantageous in certain situations—for example, with people who are extremely ill.

Changes in brain function due to brain disorders can be visualized in several ways. Magnetic resonance spectroscopy measures the concentration of specific chemical compounds in the brain that may change during specific behaviors. Functional magnetic resonance imaging (MRI) maps changes in oxygen concentration that correspond to nerve cell activity.

Positron emission tomography (PET), developed in the mid-1970s, uses computed tomography to visualize radioactive tracers (see Isotopic Tracer), radioactive substances introduced into the brain intravenously or by inhalation. PET can measure such brain functions as cerebral metabolism, blood flow and volume, oxygen use, and the formation of neurotransmitters. Single photon emission computed tomography (SPECT), developed in the 1950s and 1960s, uses radioactive tracers to visualize the circulation and volume of blood in the brain.

Brain-imaging studies have provided new insights into sensory, motor, language, and memory processes, as well as brain disorders such as epilepsy; cerebrovascular disease; Alzheimer's, Parkinson, and Huntington's diseases (see Chorea); and various mental disorders, such as schizophrenia.

VI -EVOLUTION OF THE BRAIN:

In lower vertebrates, such as fish and reptiles, the brain is often tubular and bears a striking resemblance to the early embryonic stages of the brains of more highly evolved animals. In all vertebrates, the brain is divided into three regions: the forebrain (prosencephalon), the midbrain (mesencephalon), and the hindbrain (rhomb encephalon). These three regions further subdivide into different structures, systems, nuclei, and layers.

The more highly evolved the animal, the more complex is the brain structure. Human beings have the most complex brains of all animals. Evolutionary forces have also resulted in a progressive increase in the size of the brain. In vertebrates lower than mammals, the brain is small. In meat-eating animals, particularly primates, the brain increases dramatically in size.

The cerebrum and cerebellum of higher mammals are highly convoluted in order to fit the most gray matter surface within the confines of the cranium. Such highly convoluted brains are called gyrencephalic. Many lower mammals have a smooth, or lissencephalic (―smooth head‖), cortical surface.

There is also evidence of evolutionary adaption of the brain. For example, many birds depend on an advanced visual system to identify food at great distances while in flight. Consequently, their optic lobes and cerebellum are well developed, giving them keen sight and outstanding motor coordination in flight. Rodents, on the other hand, as nocturnal animals, do not have a well-developed visual system. Instead, they rely more heavily on other sensory systems, such as a highly developed sense of smell and facial whiskers.

VII -RECENT RESEARCH:

Recent research in brain function suggests that there may be sexual differences in both brain anatomy and brain function. One study indicated that men and women may use their brains differently while thinking. Researchers used functional magnetic resonance imaging to observe which parts of the brain were activated as groups of men and women tried to determine whether sets of nonsense words rhymed. Men used only Broca's area in this task, whereas women used Broca's area plus an area on the right side of the brain

46. HEART

I - INTRODUCTION
Heart, in anatomy, hollow muscular organ that pumps blood through the body. The heart, blood, and blood vessels make up the circulatory system, which is responsible for distributing oxygen and nutrients to the body and carrying away carbon dioxide and other waste products. The heart is the circulatory system‗s power supply. It must beat ceaselessly because the body‗s tissues— especially the brain and the heart itself—depend on a constant supply of oxygen and nutrients delivered by the flowing blood. If the heart stops pumping blood for more than a few minutes, death will result.

The human heart is shaped like an upside-down pear and is located slightly to the left of center inside the chest cavity. About the size of a closed fist, the heart is made primarily of muscle tissue that contracts rhythmically to propel blood to all parts of the body. This rhythmic contraction begins in the developing embryo about three weeks after conception and continues throughout an individual‗s life. The muscle rests only for a fraction of a second between beats. Over a typical life span of 76 years, the heart will beat nearly 2.8 billion times and move 169 million liters (179 million quarts) of blood.

Since prehistoric times people have had a sense of the heart‗s vital importance. Cave paintings from 20,000 years ago depict a stylized heart inside the outline of hunted animals such as bison and elephant. The ancient Greeks believed the heart was the seat of intelligence. Others believed the heart to be the source of the soul or of the emotions—an idea that persists in popular culture and various verbal expressions, such as heartbreak, to the present day.

II - STRUCTURE OF THE HEART
The human heart has four chambers. The upper two chambers, the right and left atria, are receiving chambers for blood. The atria are sometimes known as auricles. They collect blood that pours in from veins, blood vessels that return blood to the heart. The heart's lower two chambers, the right and left ventricles, are the powerful pumping chambers. The ventricles propel blood into arteries, blood vessels that carry blood away from the heart.

A wall of tissue separates the right and left sides of the heart. Each side pumps blood through a different circuit of blood vessels: The right side of the heart pumps oxygen-poor blood to the lungs, while the left side of the heart pumps oxygen-rich blood to the body. Blood returning from a trip around the body has given up most of its oxygen and picked up carbon dioxide in the body‗s tissues. This oxygen-poor blood feeds into two large veins, the superior vena cava and inferior vena cava, which empty into the right atrium of the heart.

The right atrium conducts blood to the right ventricle, and the right ventricle pumps blood into the pulmonary artery. The pulmonary artery carries the blood to the lungs, where it picks up a fresh supply of oxygen and eliminates carbon dioxide. The blood, now oxygen-rich, returns to the heart through the pulmonary veins, which empty into the left atrium. Blood passes from the left atrium into the left ventricle, from where it is pumped out of the heart into the aorta, the body's largest artery. Smaller arteries that branch off the aorta distribute blood to various parts of the body.

A -Heart Val

Four valves within the heart prevent blood from flowing backward in the heart. The valves open easily in the direction of blood flow, but when blood pushes against the valves in the opposite direction, the valves close. Two valves, known as atrioventricular valves, are located between the atria and ventricles. The right atrioventricular valve is formed from three flaps of tissue and is called the tricuspid valve. The left atrioventricular valve has two flaps and is called the bicuspid or mitral valve. The other two heart valves are located between the ventricles and arteries. They are called semilunar valves because they each consist of three half-moon-shaped flaps of tissue. The right semilunar valve, between the right ventricle and pulmonary artery, is also called the pulmonary valve. The left semilunar valve, between the left ventricle and aorta, is also called the aortic valve.

B -Myocardium

Muscle tissue, known as myocardium or cardiac muscle, wraps around a scaffolding of tough connective tissue to form the walls of the heart's chambers. The atria, the receiving chambers of the heart, have relatively thin walls compared to the ventricles, the pumping chambers. The left ventricle has the thickest walls—nearly 1 cm (0.5 in) thick in an adult—because it must work the hardest to propel blood to the farthest reaches of the body.

C -Pericardium

A tough, double-layered sac known as the pericardium surrounds the heart. The inner layer of the pericardium, known as the epicardium, rests directly on top of the heart muscle. The outer layer of the pericardium attaches to the breastbone and other structures in the chest cavity and helps hold the heart in place. Between the two layers of the pericardium is a thin space filled with a watery fluid that helps prevent these layers from rubbing against each other when the heart beats.

D -Endocardium

The inner surfaces of the heart's chambers are lined with a thin sheet of shiny, white tissue known as the endocardium. The same type of tissue, more broadly referred to as endothelium, also lines the body's blood vessels, forming one continuous lining throughout the circulatory system. This lining helps blood flow smoothly and prevents blood clots from forming inside the circulatory system.

E -Coronary Arteries

The heart is nourished not by the blood passing through its chambers but by a specialized network of blood vessels. Known as the coronary arteries, these blood vessels encircle the heart like a crown. About 5 percent of the blood pumped to the body enters the coronary arteries, which branch from the aorta just above where it emerges from the left ventricle. Three main coronary arteries—the right, the left circumflex, and the left anterior descending—nourish different regions of the heart muscle. From these three arteries arise smaller branches that enter the muscular walls of the heart to provide a constant supply of oxygen and nutrients. Veins running through the heart muscle converge to form a large channel called the coronary sinus, which returns blood to the right atrium.

III -FUNCTION OF THE HEART

The heart's duties are much broader than simply pumping blood continuously throughout life. The heart must also respond to changes in the body's demand for oxygen. The heart works very differently during sleep, for example, than in the middle of a 5-km (3-mi) run. Moreover, the heart and the rest of the circulatory system can respond almost instantaneously to shifting situations—when a person stands up or lies down, for example, or when a person is faced with a potentially dangerous situation.

A -Cardiac Cycle

Although the right and left halves of the heart are separate, they both contract in unison, producing a single heartbeat. The sequence of events from the beginning of one heartbeat to the beginning of the next is called the cardiac cycle. The cardiac cycle has two phases: diastole, when the heart's chambers are relaxed, and systole, when the chambers contract to move blood. During the systolic phase, the atria contract first, followed by contraction of the ventricles. This sequential contraction ensures efficient movement of blood from atria to ventricles and then into the arteries. If the atria and ventricles contracted simultaneously, the heart would not be able to move as much blood with each beat.

During diastole, both atria and ventricles are relaxed, and the atrioventricular valves are open. Blood pours from the veins into the atria, and from there into the ventricles. In fact, most of the blood that enters the ventricles simply pours in during diastole. Systole then begins as the atria contract to complete the filling of the ventricles. Next, the ventricles contract, forcing blood out through the semilunar valves and into the arteries, and the atrioventricular valves close to prevent blood from flowing back into the atria. As pressure rises in the arteries, the semilunar valves snap shut to prevent blood from flowing back into the ventricles. Diastole then begins again as the heart muscle relaxes—the atria first, followed by the ventricles—and blood begins to pour into the heart once more.

A health-care professional uses an instrument known as a stethoscope to detect internal body sounds, including the sounds produced by the heart as it is beating. The characteristic heartbeat sounds are made by the valves in the heart—not by the contraction of the heart muscle itself. The sound comes from the leaflets of the valves slapping together. The closing of the atrioventricular valves, just before the ventricles contract, makes the first heart sound. The second heart sound is made when the semilunar valves snap closed. The first heart sound is generally longer and lower than the second, producing a heartbeat that sounds like lub-dup, lub-dup, lub-dup.

Blood pressure, the pressure exerted on the walls of blood vessels by the flowing blood, also varies during different phases of the cardiac cycle. Blood pressure in the arteries is higher during systole, when the ventricles are contracting, and lower during diastole, as the blood ejected during systole moves into the body's capillaries. Blood pressure is measured in millimeters (mm) of mercury using a sphygmomanometer, an instrument that consists of a pressure-recording device and an inflatable cuff that is usually placed around the upper arm. Normal blood pressure in an adult is less than 120 mm of mercury during systole, and less than 80 mm of mercury during diastole.

Blood pressure is usually noted as a ratio of systolic pressure to diastolic pressure—for example, 120/80. A person's blood pressure may increase for a short time during moments of stress or strong emotions. However, a prolonged or constant elevation of blood pressure, a condition known as hypertension, can increase a person's risk for heart attack, stroke, heart and kidney failure, and other health problems.

B -Generation of the Heartbeat

Unlike most muscles, which rely on nerve impulses to cause them to contract, heart muscle can contract of its own accord. Certain heart muscle cells have the ability to contract spontaneously, and these cells generate electrical signals that spread to the rest of the heart and cause it to contract with a regular, steady beat.

The heartbeat begins with a small group of specialized muscle cells located in the upper right-hand corner of the right atrium. This area is known as the sinoatrial (SA) node. Cells in the SA node generate their electrical signals more frequently than cells elsewhere in the heart, so the electrical signals generated by the SA node synchronize the electrical signals traveling to the rest of the heart. For this reason, the SA node is also known as the heart's pacemaker.

Impulses generated by the SA node spread rapidly throughout the atria, so that all the muscle cells of the atria contract virtually in unison. Electrical impulses cannot be conducted through the partition between the atria and ventricles, which is primarily made of fibrous connective tissue rather than muscle cells. The impulses from the SA node are carried across this connective tissue partition by a small bridge of muscle called the atrioventricular conduction system. The first part of this system is a group of cells at the lower margin of the right atrium, known as the atrioventricular (AV) node. Cells in the AV node conduct impulses relatively slowly, introducing a delay of about two-tenths of a second before an impulse reaches the ventricles. This delay allows time for the blood in the atria to empty into the ventricles before the ventricles begin contracting.

After making its way through the AV node, an impulse passes along a group of muscle fibers called the bundle of His, which span the connective tissue wall separating the atria from the ventricles. Once on the other side of that wall, the impulse spreads rapidly among the muscle cells that make up the ventricles. The impulse travels to all parts of the ventricles with the help of a network of fast-conducting fibers called Purkinje fibers. These fibers are necessary because the ventricular walls are so thick and massive.

If the impulse had to spread directly from one muscle cell to another, different parts of the ventricles would not contract together, and the heart would not pump blood efficiently. Although this complicated circuit has many steps, an electrical impulse spreads from the SA node throughout the heart in less than one second.

The journey of an electrical impulse around the heart can be traced by a machine called an electrocardiograph . This instrument consists of a recording device attached to electrodes that are placed at various points on a person's skin. The recording device measures different phases of the heartbeat and traces these patterns as peaks and valleys in a graphic image known as an electrocardiogram (ECG, sometimes known as EKG). Changes or abnormalities in the heartbeat or in the heart's rate of contraction register on the ECG, helping doctors diagnose heart problems or identify damage from a heart attack.

C -Control of the Heart Rate

In an adult, resting heart rate is normally about 70 beats per minute. However, the heart can beat up to three times faster—at more than 200 beats per minute—when a person is exercising vigorously. Younger people have faster resting heart rates than adults do. The normal heart rate is about 120 beats per minute in infants and about 100 beats per minute in young children. Many athletes, by contrast, often have relatively slow resting heart rates because physical training makes the heart stronger and enables it to pump the same amount of blood with fewer beats. An athlete's resting heart rate may be only 40 to 60 beats per minute.

Although the SA node generates the heartbeat, impulses from nerves cause the heart to speed up or slow down almost instantaneously. The nerves that affect heart rate are part of the autonomic nervous system, which directs activities of the body that are not under conscious control. The autonomic nervous system is made up of two types of nerves, sympathetic and parasympathetic fibers. These fibers come from the spinal cord or brain and deliver impulses to the SA node and other parts of the heart.

Sympathetic nerve fibers increase the heart rate. These fibers are activated in times of stress, and they play a role in the fight or flight response that prepares humans and other animals to respond to danger. In addition to fear or physical danger, exercising or experiencing a strong emotion can also activate sympathetic fibers and cause an increase in heart rate. In contrast, parasympathetic nerve fibers slow the heart rate. In the absence of nerve impulses the SA node would fire about 100 times each minute—parasympathetic fibers are responsible for slowing the heart to the normal rate of about 70 beats per minute.

Chemicals known as hormones carried in the bloodstream also influence the heart rate. Hormones generally take effect more slowly than nerve impulses. They work by attaching to receptors, proteins on the surface of heart muscle cells, to change the way the muscle cells contract. Epinephrine (also called adrenaline) is a hormone made by the adrenal glands, which are located on top of the kidneys. Released during times of stress, epinephrine increases the heart rate much as sympathetic nerve fibers do. Thyroid hormone, which regulates the body's overall metabolism, also increases the heart rate. Other chemicals—especially calcium, potassium, and sodium—can affect heart rate and rhythm.

D -Cardiac Output

To determine overall heart function, doctors measure cardiac output, the amount of blood pumped by each ventricle in one minute. Cardiac output is equal to the heart rate multiplied by the stroke volume, the amount of blood pumped by a ventricle with each beat. Stroke volume, in turn, depends on several factors: the rate at which blood returns to the heart through the veins; how vigorously the heart contracts; and the pressure of blood in the arteries, which affects how hard the heart must work to propel blood into them. Normal cardiac output in an adult is about 3 liters per minute per square meter of body surface.

An increase in either heart rate or stroke volume—or both—will increase cardiac output. During exercise, sympathetic nerve fibers increase heart rate. At the same time, stroke volume increases, primarily because venous blood returns to the heart more quickly and the heart contracts more vigorously. Many of the factors that increase heart rate also increase stroke volume. For example, impulses from sympathetic nerve fibers cause the heart to contract more vigorously as well as increasing the heart rate. The simultaneous increase in heart rate and stroke volume enables a larger and more efficient increase in cardiac output than if, say, heart rate alone increased during exercise. In a healthy adult during vigorous exercise, cardiac output can increase six-fold, to 18 liters per minute per square meter of body surface.

IV -DISEASES OF THE HEART

In the United States and many other industrialized countries, heart disease is the leading cause of death. According to the United States Centers for Disease Control and Prevention (CDC), more than 710,000 people in the United States die of heart disease each year. By far the most common type of heart disease in the United States is coronary heart disease, in which the arteries that nourish the heart become narrowed and unable to supply enough blood and oxygen to the heart muscle. However, many other problems can also affect the heart, including congenital defects (physical abnormalities that are present at birth), malfunction of the heart valves, and abnormal heart rhythms. Any type of heart disease may eventually result in heart failure, in which a weakened heart is unable to pump sufficient blood to the body.

A -Coronary Heart Disease

Coronary heart disease, the most common type of heart disease in most industrialized countries, is responsible for over 515,000 deaths in the United States yearly. It is caused by atherosclerosis, the buildup of fatty material called plaque on the inside of the coronary arteries (see Arteriosclerosis). Over the course of many years, this plaque narrows the arteries so that less blood can flow through them and less oxygen reaches the heart muscle.

The most common symptom of coronary heart disease is angina pectoris, a squeezing chest pain that may radiate to the neck, jaw, back, and left arm. Angina pectoris is a signal that blood flow to the heart muscle falls short when extra work is required from the heart muscle. An attack of angina is typically triggered by exercise or other physical exertion, or by strong emotions. Coronary heart disease can also lead to a heart attack, which usually develops when a blood clot forms at the site of a plaque and severely reduces or completely stops the flow of blood to a part of the heart. In a heart attack, also known as myocardial infarction, part of the heart muscle dies because it is deprived of oxygen. This oxygen deprivation also causes the crushing chest pain characteristic of a heart attack. Other symptoms of a heart attack include nausea, vomiting, and profuse sweating. About one-third of heart attacks are fatal, but patients who seek immediate medical attention when symptoms of a heart attack develop have a good chance of surviving.

One of the primary risk factors for coronary heart disease is the presence of a high level of a fatty substance called cholesterol in the bloodstream. High blood cholesterol is typically the result of a diet that is high in cholesterol and saturated fat, although some genetic disorders also cause the problem. Other risk factors include smoking, high blood pressure, diabetes mellitus, obesity, and a sedentary lifestyle. Coronary heart disease was once thought to affect primarily men, but this is not the case. The disease affects an equal number of men and women, although women tend to develop the disease later in life than men do.

Coronary heart disease cannot be cured, but it can often be controlled with a combination of lifestyle changes and medications. Patients with coronary heart disease are encouraged to quit smoking, exercise regularly, and eat a low-fat diet. Doctors may prescribe a drug such as lovastatin, simvastatin, or pravastatin to help lower blood cholesterol. A wide variety of medications can help relieve angina, including nitroglycerin, beta blockers, and calcium channel blockers. Doctors may recommend that some patients take a daily dose of aspirin, which helps prevent heart attacks by interfering with platelets, tiny blood cells that play a critical role in blood clotting.

In some patients, lifestyle changes and medication may not be sufficient to control angina. These patients may undergo coronary artery bypass surgery or percutaneous transluminal coronary angioplasty (PTCA) to help relieve their symptoms. In bypass surgery, a length of blood vessel is removed from elsewhere in the patient's body—usually a vein from the leg or an artery from the wrist. The surgeon sews one end to the aorta and the other end to the coronary artery, creating a conduit for blood to flow that bypasses the narrowed segment. Surgeons today commonly use an artery from the inside of the chest wall because bypasses made from this artery are very durable.

In PTCA, commonly referred to as balloon angioplasty, a deflated balloon is threaded through the patient's coronary arteries to the site of a blockage. The balloon is then inflated, crushing the plaque and restoring the normal flow of blood through the artery.

B -Congenital Defects

Each year about 25,000 babies in the United States are born with a congenital heart defect (see Birth Defects). A wide variety of heart malformations can occur. One of the most common abnormalities is a septal defect, an opening between the right and left atrium or between the right and left ventricle. In other infants, the ductus arteriosus, a fetal blood vessel that usually closes soon after birth, remains open. In babies with these abnormalities, some of the oxygen-rich blood returning from the lungs is pumped to the lungs again, placing extra strain on the right ventricle and on the blood vessels leading to and from the lung. Sometimes a portion of the aorta is abnormally narrow and unable to carry sufficient blood to the body.

This condition, called coarctation of the aorta, places extra strain on the left ventricle because it must work harder to pump blood beyond the narrow portion of the aorta. With the heart pumping harder, high blood pressure often develops in the upper body and may cause a blood vessel in the brain to burst, a complication that is often fatal. An infant may be born with several different heart defects, as in the condition known as tetralogy of Fallot. In this condition, a combination of four different heart malformations allows mixing of oxygenated and deoxygenated blood pumped by the heart. Infants with tetralogy of Fallot are often known as ―blue babies‖ because of the characteristic bluish tinge of their skin, a condition caused by lack of oxygen.

In many cases, the cause of a congenital heart defect is difficult to identify. Some defects may be due to genetic factors, while others may be the result of viral infections or exposure to certain chemicals during the early part of the mother's pregnancy. Regardless of the cause, most congenital malformations of the heart can be treated successfully with surgery, sometimes performed within a few weeks or months of birth. For example, a septal defect can be repaired with a patch made from pericardium or synthetic fabric that is sewn over the hole. An open ductus arteriosus is cut, and the pulmonary artery and aorta are stitched closed.

To correct coarctation of the aorta, a surgeon snips out the narrowed portion of the vessel and sews the normal ends together, or sews in a tube of fabric to connect the ends. Surgery for tetralogy of Fallot involves procedures to correct each part of the defect. Success rates for many of these operations are well above 90 percent, and with treatment most children with congenital heart defects live healthy, normal lives.

C -Heart Valve Malfunction

Malfunction of one of the four valves within the heart can cause problems that affect the entire circulatory system. A leaky valve does not close all the way, allowing some blood to flow backward as the heart contracts. This backward flow decreases the amount of oxygen the heart can deliver to the tissues with each beat. A stenotic valve, which is stiff and does not open fully, requires the heart to pump with increased force to propel blood through the narrowed opening. Over time, either of these problems can lead to damage of the overworked heart muscle.

Some people are born with malformed valves. Such congenital malformations may require treatment soon after birth, or they may not cause problems until a person reaches adulthood. A heart valve may also become damaged during life, due to infection, connective tissue disorders such as Marfan syndrome, hypertension, heart attack, or simply aging.

A well-known, but poorly understood, type of valve malfunction is mitral valve prolapse. In this condition, the leaflets of the mitral valve fail to close properly and bulge backward like a parachute into the left atrium. Mitral valve prolapse is the most common type of valve abnormality, affecting 5 to 10 percent of the United States population, the majority of them women. In most cases, mitral valve prolapse does not cause any problems, but in a few cases the valve's failure to close properly allows blood to leak backwards through the valve.

Another common cause of valve damage is rheumatic fever, a complication that sometimes develops after an infection with common bacteria known as streptococci. Most common in children, the illness is characterized by inflammation and pain in the joints. Connective tissue elsewhere in the body, including in the heart, heart valves, and pericardium, may also become inflamed. This inflammation can result in damage to the heart, most commonly one of the heart valves, that remains after the other symptoms of rheumatic fever have gone away.

Valve abnormalities are often detected when a health-care professional listens to the heart with a stethoscope. Abnormal valves cause extra sounds in addition to the normal sequence of two heart sounds during each heartbeat. These extra heart sounds are often known as heart murmurs, and not all of them are dangerous. In some cases, a test called echocardiography may be necessary to evaluate an abnormal valve. This test uses ultrasound waves to produce images of the inside of the heart, enabling doctors to see the shape and movement of the valves as the heart pumps.

Damaged or malformed valves can sometimes be surgically repaired. More severe valve damage may require replacement with a prosthetic valve. Some prosthetic valves are made from pig or cow valve tissue, while others are mechanical valves made from silicone and other synthetic materials.

D -Arrhythmias

Arrhythmias, or abnormal heart rhythms, arise from problems with the electrical conduction system of the heart. Arrhythmias can occur in either the atria or the ventricles. In general, ventricular arrhythmias are more serious than atrial arrhythmias because ventricular arrhythmias are more likely to affect the heart's ability to pump blood to the body.

Some people have minor arrhythmias that persist for long periods and are not dangerous—in fact, they are simply heartbeats that are normal for that particular person's heart. A temporary arrhythmia can be caused by alcohol, caffeine, or simply not getting a good night's sleep. Often, damage to the heart muscle results in a tendency to develop arrhythmias. This heart muscle damage is frequently the result of a heart attack, but can also develop for other reasons, such as after an infection or as part of a congenital defect.

Arrhythmias may involve either abnormally slow or abnormally fast rhythms.

An abnormally slow rhythm sometimes results from slower firing of impulses from the SA node itself, a condition known as sinus bradycardia. An abnormally slow heartbeat may also be due to heart block, which arises when some or all of the impulses generated by the SA node fail to be transmitted to the ventricles. Even if impulses from the atria are blocked, the ventricles continue to contract because fibers in the ventricles can generate their own rhythm. However, the rhythm they generate is slow, often only about 40 beats per minute. An abnormally slow heartbeat is dangerous if the heart does not pump enough blood to supply the brain and the rest of the body with oxygen. In this case, episodes of dizziness, lightheadedness, or fainting may occur. Episodes of fainting caused by heart block are known as Stokes-Adams attacks.

Some types of abnormally fast heart rhythms—such as atrial tachycardia, an increased rate of atrial contraction—are usually not

dangerous. Atrial fibrillation, in which the atria contract in a rapid, uncoordinated manner, may reduce the pumping efficiency of the heart. In a person with an otherwise healthy heart, this may not be dangerous, but in a person with other heart disease the reduced pumping efficiency may lead to heart failure or stroke.
By far the most dangerous type of rapid arrhythmia is ventricular fibrillation, in which ventricular contractions are rapid and chaotic. Fibrillation prevents the ventricles from pumping blood efficiently, and can lead to death within minutes. Ventricular fibrillation can be reversed with an electrical defibrillator, a device that delivers a shock to the heart. The shock briefly stops the heart from beating, and when the heartbeat starts again the SA node is usually able to resume a normal beat.

Most often, arrhythmias can be diagnosed with the use of an ECG. Some arrhythmias do not require treatment. Others may be controlled with medications such as digitalis, propanolol, or disopyramide. Patients with heart block or several other types of arrhythmias may have an artificial pacemaker implanted in their chest. This small, battery-powered electronic device delivers regular electrical impulses to the heart through wires attached to different parts of the heart muscle. Another type of implantable device, a miniature defibrillator, is used in some patients at risk for serious ventricular arrhythmias. This device works much like the larger defibrillator used by paramedics and in the emergency room, delivering an electric shock to reset the heart when an abnormal rhythm is detected.

E -Other Forms of Heart Disease
In addition to the relatively common heart diseases described above, a wide variety of other diseases can also affect the heart. These include tumors, heart damage from other diseases such as syphilis and tuberculosis, and inflammation of the heart muscle, pericardium, or endocardium.
Myocarditis, or inflammation of the heart muscle, was commonly caused by rheumatic fever in the past. Today, many cases are due to a viral infection or their cause cannot be identified. Sometimes myocarditis simply goes away on its own. In a minority of patients, who often suffer repeated episodes of inflammation, myocarditis leads to permanent damage of the heart muscle, reducing the heart's ability to pump blood and making it prone to developing abnormal rhythms.

Cardiomyopathy encompasses any condition that damages and weakens the heart muscle. Scientists believe that viral infections cause many cases of cardiomyopathy. Other causes include vitamin B deficiency, rheumatic fever, underactivity of the thyroid gland, and a genetic disease called hemochromatosis in which iron builds up in the heart muscle cells. Some types of cardiomyopathy can be controlled with medication, but others lead to progressive weakening of the heart muscle and sometimes result in heart failure.

In pericarditis, the most common disorder of the pericardium, the saclike membrane around the heart becomes inflamed. Pericarditis is most commonly caused by a viral infection, but may also be due to arthritis or an autoimmune disease such as systemic lupus erythematosus. It may be a complication of late-stage kidney disease, lung cancer, or lymphoma; it may be a side effect of radiation therapy or certain drugs. Pericarditis sometimes goes away without treatment, but it is often treated with antiinflammatory drugs. It usually causes no permanent damage to the heart. If too much fluid builds up around the heart during an attack of pericarditis, the fluid may need to be drained with a long needle or in a surgical procedure. Patients who suffer repeated episodes of pericarditis may have the pericardium surgically removed.

Endocarditis is an infection of the inner lining of the heart, but damage from such an infection usually affects only the heart valves. Endocarditis often develops when bacteria from elsewhere in the body enter the bloodstream, settle on the flaps of one of the heart valves, and begin to grow there. The infection can be treated with antibiotics, but if untreated, endocarditis is often fatal. People with congenital heart defects, valve damage due to rheumatic fever, or other valve problems are at greatest risk for developing endocarditis. They often take antibiotics as a preventive measure before undergoing dental surgery or certain other types of surgery that can allow bacteria into the bloodstream. Intravenous drug users who share needles are another population at risk for endocarditis. People who use unclean needles, which allow bacteria into the bloodstream, frequently develop valve damage.

F -Heart Failure
The final stage in almost any type of heart disease is heart failure, also known as congestive heart failure, in which the heart muscle weakens and is unable to pump enough blood to the body. In the early stages of heart failure, the muscle may enlarge in an attempt to contract more vigorously, but after a time this enlargement of the muscle simply makes the heart inefficient and unable to deliver enough blood to the tissues. In response to this shortfall, the kidneys conserve water in an attempt to increase blood volume, and the heart is stimulated to pump harder. Eventually excess fluid seeps through the walls of tiny blood vessels and into the tissues. Fluid may collect in the lungs, making breathing difficult, especially when a patient is lying down at night. Many patients with heart failure must sleep propped up on pillows to be able to breathe. Fluid may also build up in the ankles, legs, or abdomen. In the later stages of heart failure, any type of physical activity becomes next to impossible.

Almost any condition that overworks or damages the heart muscle can eventually result in heart failure. The most common cause of heart failure is coronary heart disease. Heart failure may develop when the death of heart muscle in a heart attack leaves the heart with less strength to pump blood, or simply as a result of long-term oxygen deprivation due to narrowed coronary arteries. Hypertension or malfunctioning valves that force the heart to work harder over extended periods of time may also lead to heart failure. Viral or bacterial infections, alcohol abuse, and certain chemicals (including some lifesaving drugs used in cancer chemotherapy), can all damage the heart muscle and result in heart failure.

Despite its ominous name, heart failure can sometimes be reversed and can often be effectively treated for long periods with a combination of drugs. About 4.6 million people with heart failure are alive in the United States today. Medications such as digitalis are often prescribed to increase the heart's pumping efficiency, while beta blockers may be used to decrease the heart's workload. Drugs known as vasodilators relax the arteries and veins so that blood encounters less resistance as it flows. Diuretics stimulate the kidneys to excrete excess fluid.

A last resort in the treatment of heart failure is heart transplantation, in which a patient's diseased heart is replaced with a healthy heart from a person who has died of other causes . Heart transplantation enables some patients with heart failure to lead active, healthy lives once again. However, a person who has received a heart transplant must take medications to suppress the immune system for the rest of his or her life in order to prevent rejection of the new heart. These drugs can have serious side effects, making a person more vulnerable to infections and certain types of cancer.
The first heart transplant was performed in 1967 by South African surgeon Christiaan Barnard. However, the procedure did not become widespread until the early 1980s, when the immune-suppressing drug cyclosporine became available. This drug helps prevent rejection without making patients as vulnerable to infection as they had been with older immune-suppressing drugs. About 3,500 heart transplants are performed worldwide each year, about 2,500 of them in the United States. Today, about 83 percent of heart transplant recipients survive at least one year, and 71 percent survive for four years.

A shortage of donor hearts is the main limitation on the number of transplants performed today. Some scientists are looking for alternatives to transplantation that would help alleviate this shortage of donor hearts. One possibility is to replace a human heart with a mechanical one. A permanent artificial heart was first implanted in a patient in 1982. Artificial hearts have been used experimentally with mixed success. They are not widely used today because of the risk of infection and bleeding and concerns about their reliability. In addition, the synthetic materials used to fashion artificial hearts can cause blood clots to form in the heart. These blood clots may travel to a vessel in the neck or head, resulting in a stroke. Perhaps a more promising option is the left ventricular assist device (LVAD). This device is implanted inside a person's chest or abdomen to help the patient's own heart pump blood. LVADs are used in many people waiting for heart transplants, and could one day become a permanent alternative to transplantation.

Some scientists are working to develop xenotransplantation, in which a patient's diseased heart would be replaced with a heart from a pig or another species. However, this strategy still requires a great deal of research to prevent the human immune system from rejecting a heart from a different species. Some experts have also raised concerns about the transmission of harmful viruses from other species to humans as a result of xenotransplantation.

V -HISTORY OF HEART RESEARCH

Scientific knowledge of the heart dates back almost as far as the beginnings of recorded history. The Egyptian physician Imhotep made observations on the pulse during the 2600s BC. During the 300s BC the Greek physician Hippocrates studied and wrote about various signs and symptoms of heart disease, and the Greek philosopher Aristotle described the beating heart of a chick embryo. Among the first people to investigate and write about the anatomy of the heart was another Greek physician, Erasistratus, around 250 BC. Erasistratus described the appearance of the heart and the four valves inside it. Although he correctly deduced that the valves prevent blood from flowing backward in the heart, he did not understand that the heart was a pump. Galen, a Greek-born Roman physician, also wrote about the heart during the second century AD. He recognized that the heart was made of muscle, but he believed that the liver was responsible for the movement of blood through the body.

Heart research did not greatly expand until the Renaissance in Europe (14th century to 16th century). During that era, scientists began to connect the heart's structure with its function. During the early 16th century the Spanish physician and theologian Michael Servetus described how blood passes through the four chambers of the heart and picks up oxygen in the lungs. Perhaps the most significant contributions were made by English physician William Harvey, who discovered the circulation of blood in 1628. Harvey was the first to realize that the heart is a pump responsible for the movement of blood through the body. His work revealed how the heart works with the blood and blood vessels to nourish the body, establishing the concept of the circulatory system.

The 20th century witnessed extraordinary advances in the diagnosis of heart diseases, corrective surgeries, and other forms of treatment for heart problems. Many doctors had become interested in measuring the pulse and abnormal heartbeats. This line of research culminated in the 1902 invention of the electrocardiograph by Dutch physiologist Willem Einthoven, who received the Nobel Prize for this work in 1924. Another major advance in diagnosis was cardiac catheterization, which was pioneered in 1929 by German physician Werner Forssmann. After performing experiments on animals, Forssmann inserted a catheter through a vein in his arm and into his own heart—a stunt for which he was fired from his job. Two American physicians, André Cournand and Dickinson Richards, later continued research on catheterization, and the technique became commonly used during the 1940s. The three scientists received the Nobel Prize in 1956 for their work.

At the beginning of the 20th century, most doctors believed that surgery on the heart would always remain impossible, as the heart was thought to be an extremely delicate organ. Most of the first heart operations were done in life-or-death trauma situations. American physician L. L. Hill performed the first successful heart surgery in the United States in 1902, sewing up a stab wound in the left ventricle of an 8-year-old boy. The next year, French surgeon Marin Théodore Tuffier removed a bullet from a patient's left atrium.

Surgery to correct some congenital defects involving blood vessels also helped lay the foundations for surgery on the heart itself. In 1938 American surgeon Robert Gross performed the first successful surgery to treat an open ductus arteriosus, tying the vessel closed with thread. In 1944 Gross and Swedish surgeon Clarence Crafoord each performed successful surgery for coarctation of the aorta. The same year, American surgeon Alfred Blalock and surgical assistant Vivien Thomas performed the first successful operation to correct tetralogy of Fallot. But the greatest leap forward came in 1953, when American physician John Gibbon introduced the heart-lung machine, a device to oxygenate and pump blood during surgery on the heart. This invention made open-heart surgery—with the heart stopped for the duration of the operation—possible. It led to now-routine surgical techniques such as valve replacement, correction of congenital defects, and bypass surgery.

The rapid pace of scientific discovery during the 20th century has also led to many nonsurgical treatments for diseases of the heart. The introduction of antibiotics to treat bacterial infections greatly reduced sickness and deaths due to heart disease from rheumatic fever, endocarditis, and other infections involving the heart, although these infections remain a significant threat in many developing nations. Many effective drugs to control hypertension, reduce cholesterol, relieve angina, limit damage from heart attacks, and treat other forms of heart disease have also been developed. Advances in electronics led to implantable pacemakers in 1959 and implantable defibrillators in 1982.

VI -HEARTS IN OTHER ANIMALS
Among different groups of animals, hearts vary greatly in size and complexity. In insects, the heart is a hollow bulb with muscular walls that contract to push blood into an artery. Many insects have several such hearts arranged along the length of the artery. When the artery ends, blood percolates among the cells of the insect's body, eventually making its way back to the heart. In an insect, blood may take as long as an hour to complete a trip around the body.

In earthworms and other segmented worms, known as annelids, blood flows toward the back of the body through the ventral blood vessel and toward the front of the body through the dorsal blood vessel. Five pairs of hearts, or aortic arches, help pump blood. The hearts are actually segments of the dorsal blood vessel and are similar in structure to those of insects.

In vertebrates, or animals with a backbone, the heart is a separate, specialized organ rather than simply a segment of a blood vessel. In fish, the heart has two chambers: an atrium (receiving chamber) and a ventricle (pumping chamber). Oxygen-depleted blood returning from the fish's body empties into the atrium, which pumps blood into the ventricle. The ventricle then pumps the blood to the gills, the respiratory organs of fish. In the gills, the blood picks up oxygen from the water and gets rid of carbon dioxide. The freshly oxygenated blood leaves the gills and travels to various parts of the body. In fish, as in humans, blood passes through the respiratory organs before it is distributed to the body. Unlike in humans, the blood does not return to the heart between visiting the respiratory organs and being distributed to the tissues. Without the added force from a second trip through the heart, blood flows relatively slowly in fish compared to humans and other mammals. However, this sluggish flow is enough to supply the fish's relatively low oxygen demand.

As vertebrates moved from life in the sea to life on land, they evolved lungs as new respiratory organs for breathing. At the same time, they became more active and developed greater energy requirements. Animals use oxygen to release energy from food molecules in a process called cellular respiration, so land-dwelling vertebrates also developed a greater requirement for oxygen. These changes, in turn, led to changes in the structure of the heart and circulatory system. Amphibians and most reptiles have a heart with three chambers—two atria and a single ventricle. These animals also have separate circuits of blood vessels for oxygenating blood and delivering it to the body.

Deoxygenated blood returning from the body empties into the right atrium. From there, blood is conducted to the ventricle and is then pumped to the lungs. After picking up oxygen and getting rid of carbon dioxide in the lungs, blood returns to the heart and empties into the left atrium. The blood then enters the ventricle a second time and is pumped out to the body. The second trip through the heart keeps blood pressure strong and blood flow rapid as blood is pumped to the tissues, helping the blood deliver oxygen more efficiently.

The three-chambered heart of amphibians and reptiles also creates an opportunity for blood to mix in the ventricle which pumps both oxygenated and deoxygenated blood with each beat. While in birds and mammals this would be deadly, scientists now understand that a three-chambered heart is actually advantageous for amphibians and reptiles. These animals do not breathe constantly—for example, amphibians absorb oxygen through their skin when they are underwater—and the three-chambered heart enables them to adjust the proportions of blood flowing to the body and the lungs depending on whether the animal is breathing or not. The three-chambered heart actually results in more efficient oxygen delivery for amphibians and reptiles.

Birds and mammals have high-energy requirements even by vertebrate standards, and a corresponding high demand for oxygen. Their hearts have four chambers—two atria and two ventricles—resulting in a complete separation of oxygenated and deoxygenated blood and highly efficient delivery of oxygen to the tissues. Small mammals have more rapid heart rates than large mammals because they have the highest energy needs. The resting heart rate of a mouse is 500 to 600 beats per minute, while that of an elephant is 30 beats per minute. Blood pressure also varies among different mammal species. Blood pressure in a giraffe's aorta is about 220 mm of mercury when the animal is standing. This pressure would be dangerously high in a human, but is necessary in a giraffe to lift blood up the animal's long neck to its brain.

Although other groups of vertebrates have hearts with a different structure than those of humans, they are still sufficiently similar that scientists can learn about the human heart from other animals. Scientists use a transparent fish, the zebra fish, to learn how the heart and the blood vessels that connect to it form before birth. Fish embryos are exposed to chemicals known to cause congenital heart defects, and scientists look for resulting genetic changes. Researchers hope that these studies will help us understand why congenital heart malformations occur, and perhaps one day prevent these birth defects

The human heart:

The human heart is a hollow, pear-shaped organ about the size of a fist. The heart is made of muscle that rhythmically contracts, or beats, pumping blood throughout the body. Oxygen-poor blood from the body enters the heart from two large blood vessels, the inferior vena cava and the superior vena cava, and collects in the right atrium. When the atrium fills, it contracts, and blood passes through the tricuspid valve into the right ventricle. When the ventricle becomes full, it starts to contract, and the tricuspid valve closes to prevent blood from moving back into the atrium.

As the right ventricle contracts, it forces blood into the pulmonary artery, which carries blood to the lungs to pick up fresh oxygen. When blood exits the right ventricle, the ventricle relaxes and the pulmonary valve shuts, preventing blood from passing back into the ventricle. Blood returning from the lungs to the heart collects in the left atrium. When this chamber contracts, blood flows through the mitral valve into the left ventricle. The left ventricle fills and begins to contract, and the mitral valve between the two chambers closes. In the final phase of blood flow through the heart, the left ventricle contracts and forces blood into the aorta. After the blood in the left ventricle has been forced out, the ventricle begins to relax, and the aortic valve at the opening of the aorta closes.

Valves

Thin, fibrous flaps called valves lie at the opening of the heart's pulmonary artery and aorta. Valves are also present between each atrium and ventricle of the heart. Valves prevent blood from flowing backward in the heart. In this illustration of the pulmonary valve, as the heart contracts, blood pressure builds and pushes blood up against the pulmonary valve, forcing it to open. As the heart relaxes between one beat and the next, blood pressure falls. Blood flows back from the pulmonary artery, forcing the pulmonary valve to close, and preventing backflow of blood

47. TISSUES

I -INTRODUCTION
Tissue, group of associated, similarly structured cells that perform specialized functions for the survival of the organism. Animal tissues, to which this article is limited, take their first form when the blastula cells, arising from the fertilized ovum, differentiate into three germ layers: the ectoderm, mesoderm, and endoderm. Through further cell differentiation, or histogenesis, groups of cells grow into more specialized units to form organs made up, usually, of several tissues of similarly performing cells. Animal tissues are classified into four main groups.

II -EPITHELIAL TISSUES
These tissues include the skin and the inner surfaces of the body, such as those of the lungs, stomach, intestines, and blood vessels. Because its primary function is to protect the body from injury and infection, epithelium is made up of tightly packed cells with little intercellular substance between them.

About 12 kinds of epithelial tissue occur. One kind is stratified squamous tissue found in the skin and the linings of the esophagus and vagina. It is made up of thin layers of flat, scalelike cells that form rapidly above the blood capillaries and are pushed toward the tissue surface, where they die and are shed. Another is simple columnar epithelium, which lines the digestive system from the stomach to the anus; these cells stand upright and not only control the absorption of nutrients but also secrete mucus through individual goblet cells. Glands are formed by the inward growth of epithelium—for example, the sweat glands of the skin and the gastric glands of the stomach. Outward growth results in hair, nails, and other structures.

III -CONNECTIVE TISSUES
These tissues, which support and hold parts of the body together, comprise the fibrous and elastic connective tissues, the adipose (fatty) tissues, and cartilage and bone. In contrast to epithelium, the cells of these tissues are widely separated from one another, with a large amount of intercellular substance between them. The cells of fibrous tissue, found throughout the body, connect to one another by an irregular network of strands, forming a soft, cushiony layer that also supports blood vessels, nerves, and other organs. Adipose tissue has a similar function, except that its fibroblasts also contain and store fat. Elastic tissue, found in ligaments, the trachea, and the arterial walls, stretches and contracts again with each pulse beat. In the human embryo, the fibroblast cells that originally secreted collagen for the formation of fibrous tissue later change to secrete a different form of protein called chondrion, for the formation of cartilage; some cartilage later becomes calcified by the action of osteoblasts to form bones. Blood and lymph are also often considered connective tissues.

IV -MUSCLE TISSUES
These tissues, which contract and relax, comprise the striated, smooth, and cardiac muscles. Striated muscles, also called skeletal or voluntary muscles, include those that are activated by the somatic, or voluntary, nervous system. They are joined together without cell walls and have several nuclei. The smooth, or involuntary muscles, which are activated by the autonomic nervous system, are found in the internal organs and consist of simple sheets of cells. Cardiac muscles, which have characteristics of both striated and smooth muscles, are joined together in a vast network of interlacing cells and muscle sheaths.

V -NERVE TISSUES
These highly complex groups of cells, called ganglia, transfer information from one part of the body to another. Each neuron, or nerve cell, consists of a cell body with branching dendrites and one long fiber, or axon. The dendrites connect one neuron to another; the axon transmits impulses to an organ or collects impulses from a sensory organ

48. Epithelial Cell

A color-enhanced microscopic photograph reveals the distribution of structures and substances in epithelial cells isolated from the pancreas. The red areas correspond to deoxyribonucleic acid, the blue to microtubules, and the green to actins. The cells secrete bicarbonate which neutralizes acid

49. ORIGINS OF MODERN HUMANS

Multiregional or Out of Africa?

Around 30,000 years ago humans were anatomically and behaviorally similar throughout the world.

One of the most hotly debated issues in paleoanthropology (the study of human origins) focuses on the origins of modern humans, Homo sapiens.9,10,3,6,13,15,14 Roughly 100,000 years ago, the Old World was occupied by a morphologically diverse group of hominids. In Africa and the Middle East there was Homo sapiens; in Asia, Homo erectus; and in Europe, Homo neanderthalensis. However, by 30,000 years ago this taxonomic diversity vanished and humans everywhere had evolved into the anatomically and behaviorally modern form. The nature of this transformation is the focus of great deliberation between two schools of thought: one that stresses multiregional continuity and the other that suggests a single origin for modern humans.

Multiregional theory: Homo erectus left Africa 2 mya to become Homo sapiens in different parts of the world.

Understanding the issue

The Multiregional Continuity Model15 contends that after Homo erectus left Africa and dispersed into other portions of the Old World, regional populations slowly evolved into modern humans. This model contains the following components:

• some level of gene flow between geographically separated populations prevented speciation, after the dispersal

• All living humans derive from the species Homo erectus that left Africa nearly two million-years-ago

• Natural selection in regional populations, ever since their original dispersal, is responsible for the regional variants (sometimes called races) we see today the emergence of Homo sapiens was not restricted to any one area, but was a phenomenon that occurred throughout the entire geographic range where humans lived.

Out of Africa theory:
Homo sapiens arose in Africa and migrated to other parts of the world to replace other hominid species, including Homo erectus

In contrast, the Out of Africa Model13 asserts that modern humans evolved relatively recently in Africa, migrated into Eurasia and replaced all populations which had descended from Homo erectus. Critical to this model are the following tenets:

• after Homo erectus migrated out of Africa the different populations became reproductively isolated, evolving independently, and in some cases like the Neanderthals, into separate species

• Homo sapiens arose in one place, probably Africa (geographically this includes the Middle East)

• Homo sapiens ultimately migrated out of Africa and replaced all other human populations, without interbreeding

• Modern human variation is a relatively recent phenomenon
The multiregional view posits that genes from all human populations of the Old World flowed between different regions and by mixing together, contributed to what we see today as fully modern humans. The replacement hypothesis suggests that the genes in fully modern humans all came out of Africa. As these peoples migrated they replaced all other human populations with little or no interbreeding. To understand this controversy, the anatomical, archaeological, and genetic evidence needs to be evaluated.

Anatomical evidence

Sometime prior to 1 million years ago early hominids, sometimes referred to as Homo ergaster, exited Africa and dispersed into other parts of the Old World. Living in disparate geographical areas their morphology became diversified through the processes of genetic drift and natural selection.

• In Asia these hominids evolved into Peking Man and Java Man, collectively referred to as Homo erectus.

• In Europe and western Asia they evolved into the Neanderthals.
Neanderthals lived in quasi isolation in Europe during a long, relatively cool period that even included glaciations. Neanderthals are distinguished by a unique set of anatomical features, including:

• A large, long, low cranial vault with a well-developed double-arched browridge

• A massive facial skeleton with a very projecting mid-face, backward sloping cheeks, and large nasal aperture, with large nasal sinuses

• An oddly shaped occipital region of the skull with a bulge or bun

• Molars with enlarged pulp chambers, and large, often very heavily worn incisors

• A mandible lacking a chin and possessing a large gap behind the last molar

• A massive thorax, and relatively short forearms and lower legs

• Although short in stature they possessed robustly built skeletons with thick walled limb bones long clavicles and very wide scapulas.

Homo sapiens is a separate species from Neanderthals and other hominids

By 130,000 years ago, following a prolonged period of independent evolution in Europe, Neanderthals were so anatomically distinct that they are best classified as a separate species -- Homo neanderthalensis. This is a classic example of geographic isolation leading to a speciation event. In contrast, at roughly the same time, in Africa, a body plans essentially like our own had appeared. While these early Homo sapiens were anatomically modern they were not behaviorally modern. It is significant that modern anatomy evolved prior to modern behavior. These early sapiens were characterized by:

• a cranial vault with a vertical forehead, rounded occipital and reduced brow ridge
• a reduced facial skeleton lacking a projecting mid-face
• a lower jaw sporting a chin
• a more modern, less robustly built skeleton.

Hence, the anatomical and paleogeographic evidence suggests that Neanderthals and early modern humans had been isolated from one another and were evolving separately into two distinct species.

Homo sapiens exhibited technological skills around 50,000 years ago.

Archaeological evidence

Very interestingly, while Neanderthals and early Homo sapiens were distinguished from one another by a suite of obvious anatomical features, archaeologically they were very similar. Hominids of the Middle Stone Age of Africa (H. sapiens) and their contemporary Middle Paleolithic Neanderthals of Europe had artifact assemblages characterized as follows:

• little variation in stone tool types, with a preponderance of flake tools that are difficult to sort into discrete categories
• over long periods of time and wide geographical distances there was general similarity in tool kits
• a virtual lack of tools fashioned out of bone, antler or ivory
• burials lacked grave goods and signs of ritual or ceremony
• hunting was usually limited to less dangerous species and evidence for fishing is absent
• population densities were apparently low
• no evidence of living structures exist and fireplaces are rudimentary
• evidence for art or decoration is also lacking

The archaeological picture changed dramatically around 40-50,000 years ago with the appearance of behaviorally modern humans. This was an abrupt and dramatic change in subsistence patterns, tools and symbolic expression. The stunning change in cultural adaptation was not merely a quantitative one, but one that represented a significant departure from all earlier human behavior, reflecting a major qualitative transformation. It was literally a "creative explosion" which exhibited the "technological ingenuity, social formations, and ideological complexity of historic hunter-gatherers."7 This human revolution is precisely what made us who we are today.

The appearance of fully modern behavior apparently occurred in Africa earlier than anywhere else in the Old World, but spread very quickly, due to population movements into other geographical regions. The Upper Paleolithic lifestyle, as it was called, was based essentially on hunting and gathering. So successful was this cultural adaptation that until roughly 11,000 years ago, hominids worldwide were subsisting essentially as hunter-gatherers.
In the Upper Paleolithic of Eurasia, or the Late Stone Age as it is called in Africa, the archaeological signature stands in strong contrast to that of the Middle Paleolithic/Middle Stone Age. It was characterized by significant innovation:

•a remarkable diversity in stone tool types
•tool types showed significant change over time and space
•artifacts were regularly fashioned out of bone, antler and ivory, in addition to stone
•stone artifacts were made primarily on blades and were easily classified into discrete categories, presumably reflecting specialized use
•burials were accompanied by ritual or ceremony and contained a rich diversity of grave goods
•living structures and well-designed fireplaces were constructed
•hunting of dangerous animal species and fishing occurred regularly
higher population densities
•abundant and elaborate art as well as items of personal adornment were widespread
•raw materials such as flint and shells were traded over some distances

Homo sapiens of the Upper Paleolithic/Late Stone Age was quintessentially modern in appearance and behavior. Precisely how this transformation occurred is not well understood, but it apparently was restricted to Homo sapiens and did not occur in Neanderthals. Some archaeologists invoke a behavioral explanation for the change. For example, Soffer11 suggests that changes in social relations, such as development of the nuclear family, played a key role in bringing about the transformation.

Social or biological changes may account for "smarter" hominids

Klein7, on the other hand, proffers the notion that it was probably a biological change brought about by mutations that played the key role in the emergence of behaviorally modern humans. His biologically based explanation implies that a major neural reorganization of the brain resulted in a significant enhancement in the manner in which the brain processed information. This is a difficult hypothesis to test since brains do not fossilize. But it is significant that no changes are seen in the shape of the skulls between earlier and later Homo sapiens. It can only be surmised from the archaeological record, which contains abundant evidence for ritual and art, that these Upper Paleolithic/Late Stone Age peoples possessed language abilities equivalent to our own. For many anthropologists this represents the final evolutionary leap to full modernity.

Shortly after fully modern humans entered Europe, roughly 40,000 years ago, the Neanderthals began a fairly rapid decline, culminating in their disappearance roughly 30,000 years ago. Neanderthals were apparently no match for the technologically advanced fully modern humans who invaded Europe and evidence for interbreeding of these two types of hominids is equivocal.

Africans display higher genetic variation than other populations, supporting the idea that they were the first modern humans.

Genetic evidence

Investigation of the patterns of genetic variation in modern human populations supports the view that the origin of Homo sapiens is the result of a recent event that is consistent with the Out of Africa Model.

• Studies of contemporary DNA, especially mitochondrial DNA (mtDNA) which occurs only in the cellular organelles called mitochondria, reveal that humans are astonishingly homogeneous, with relatively little genetic variation.1,5

• The high degree of similarity between human populations stands in strong contrast to the condition seen in our closest living relatives, the chimpanzees.2 In fact, there is significantly more genetic variation between two individual chimpanzees drawn from the same population than there is between two humans drawn randomly from a single population. Furthermore, genetic variation between populations of chimpanzees is enormously greater than differences between European, Asian and African human populations.

• In support of an African origin for Homo sapiens the work of Cann and Wilson1 has demonstrated that the highest level of genetic variation in mtDNA occurs in African populations. This implies that Homo sapiens arose first in Africa and has therefore had a longer period of time to accumulate genetic diversity. Using the genetic distance between African populations and others as a measure of time, they furthermore suggested that Homo sapiens arose between 100,000 and 400,000 years ago in Africa.

• The low amount of genetic variation in modern human populations suggests that our origins may reflect a relatively small founding population for Homo sapiens. Analysis of mtDNA by Rogers and Harpending12 supports the view that a small population of Homo sapiens, numbering perhaps only 10,000 to 50,000 people, left Africa somewhere between 50,000 and 100,000 years ago.

• Scientists recently succeeded in extracting DNA from several Neanderthal skeletons.8 After careful analysis of particularly the mtDNA, but now also some nuclear DNA, it is apparent that Neanderthal DNA is very distinct from our own. In assessing the degree of difference between DNA in Neanderthals and modern humans, the authors suggest that these two lineages have been separated for more than 400,000 years.
Although in its infancy, such genetic studies support the view that Neanderthals did not interbreed with Homo sapiens who migrated into Europe. It is, therefore, highly likely that modern humans do not carry Neanderthal genes in their DNA.

Neanderthals and modern humans coexisted in some parts of the world for thousands of years. Neanderthals probably did not breed with modern humans but they borrowed some of their tools and skills.

Additional considerations

The chronology in the Middle East does not support the Multiregional Model where Neanderthals and anatomically modern humans overlapped for a long period of time.

• Cave sites in Israel, most notably Qafzeh and Skhul date to nearly 100,000 years and contain skeletons of anatomically modern humans. Furthermore, Neanderthal remains are known from sites such as the 110,000-year-old Tabun cave, which predates the earliest Homo sapiens by about 10,000 years in the region.

• The presence of Neanderthals at two other caves in Israel, Amud and Kebara, dated to roughly 55,000 years means that Neanderthals and Homo sapiens overlapped in this region for at least 55,000 years. Therefore, if Homo sapiens were in this region for some 55,000 years prior to the disappearance of the Neanderthals, there is no reason to assume that Neanderthals evolved into modern humans.

• Archaeological evidence from Europe suggests that Neanderthals may have survived in the Iberian Peninsula until perhaps as recently as 30,000 to 35,000 years ago. Fully modern humans first appear in Europe at around 35,000-40,000 years ago, bringing with them an Upper Paleolithic tool tradition referred to as the Aurignacian. Hence, Neanderthals and fully modern humans may have overlapped for as much as 10,000 years in Europe. Again, with fully modern humans on the scene, it is not necessary to have Neanderthals evolve into modern humans, further bolstering the view that humans replaced Neanderthals.

• The situation in southern France is, however, not quite as clear. Here, at several sites, dating to roughly 40,000 years there is evidence of an archaeological industry called the Châtelperronian that contains elements of Middle and Upper Paleolithic artifacts. Hominids from these sites are clearly Neanderthals, sparking speculation that the Châtelperronian is an example of Neanderthals mimicking the culture of modern humans. The lack of anatomical intermediates at these sites, suggests that if Neanderthals did encounter and borrow some technology from Homo sapiens, they did not interbreed.

• A potential 24,500-year-old Neanderthal/sapiens hybrid was announced from the site of Lagar Velho, Portugal.4 This 4-year-old has a short, squat body like a Neanderthal, but possesses an anatomically modern skull. There are a number of problems with interpreting this find as a Neanderthal/sapiens hybrid.14 First of all, as a hybrid it should have a mixture of traits throughout its body and not possess the body of a Neanderthal and skull of a modern human. For example, if we look at hybrids of lions and tigers they do not possess the head of one species and the body of the other, but exhibit a morphological mixture of the two species. Secondly, and more importantly, acceptance of this specimen as a hybrid would suggest that Neanderthal traits had been retained for some 6,000 to 10,000 years after Neanderthals went extinct, which is highly unlikely. This is theoretically unlikely since Neanderthal traits would have been genetically swamped by the Homo sapiens genes over such a protracted period of time.

• Proponents of the Multiregional Model, such as Milford Wolpoff, cite evidence in Asia of regional continuity. They see an evolutionary link between ancient Homo erectus in Java right through to Australian aborigines. A possible problem with this view is that recent dating of late surviving Homo erectus in Indonesia suggests that they survived here until 50,000 years ago, which is potentially when fully modern humans may have arrived in the region from Africa.

China may contain the best evidence for supporting the Multiregional Model. Here there are discoveries of a couple of skulls dated to roughly 100,000 years ago that seem to possess a mixture of classic Homo erectus and Homo sapiens traits. Better geological dating and more complete specimens are needed to more fully assess this possibility.

Conclusion
For the moment, the majority of anatomical, archaeological and genetic evidence gives credence to the view that fully modern humans are a relatively recent evolutionary phenomenon. The current best explanation for the beginning of modern humans is the Out of Africa Model that postulates a single, African origin for Homo sapiens. The major neurological and cultural innovations that characterized the appearance of fully modern humans has proven to be remarkably successful, culminating in our dominance of the planet at the expense of all earlier hominid populations.

50. PEST CONTROL

I -INTRODUCTION
Pest Control, any of a wide range of environmental interventions that have as their objective the reduction to acceptable levels of insect pests, plant pathogens, and weed populations. Specific control techniques include chemical, physical, and biological mechanisms. Despite all the control efforts used, pests annually destroy about 35 percent of all crops worldwide. Even after food is harvested, insects, microorganisms, rodents, and birds inflict a further 10 to 20 percent loss, bringing the total destruction to about 40 or 50 percent. With so many areas of the world facing serious food shortages, researchers seek to reduce this loss by improving pest control.

II -CHEMICAL CONTROLS
The chemical agents called pesticides include herbicides (for weed control), insecticides, and fungicides. More than half the pesticides used in the United States are herbicides that control weeds. The United States Department of Agriculture (USDA) estimates indicate that 86 percent of U.S. agricultural land areas are treated with herbicides, 18 percent with insecticides, and 3 percent with fungicides. The amount of pesticide used on different crops also varies. For example, in the United States, about 67 percent of the insecticides used in agriculture are applied to two crops, cotton and corn; about 70 percent of the herbicides are applied to corn and soybeans, and most of the fungicides are applied to fruit and vegetable crops.

Most of the insecticides now applied are long-lasting synthetic compounds that affect the nervous system of insects on contact. Among the most effective are the chlorinated hydrocarbons DDT, chlordane, and toxaphene, although agricultural use of DDT has been banned in the United States since 1973. Others, the organophosphate insecticides, include malathion, parathion, and dimethoate. Among the most effective herbicides are the compounds of 2,4-D (2,4-dichlorophenoxyacetic acid), only a few kilograms of which are required per hectare to kill broad-leaved weeds while leaving grains unaffected.

Agricultural pesticides prevent a monetary loss of about $9 billion each year in the United States. For every $1 invested in pesticides, the American farmer gets about $4 in return. These benefits, however, must be weighed against the costs to society of using pesticides, as seen in the banning of ethylene dibromide in the early 1980s. These costs include human poisonings, fish kills, honeybee poisonings, and the contamination of livestock products. The environmental and social costs of pesticide use in the United States have been estimated to be at least $1 billion each year. Thus, although pesticides are valuable for agriculture, they also can cause serious harm.

Indeed, the question may be asked—what would crop losses be if insecticides were not used in the United States, and readily available nonchemical controls were substituted? The best estimate is that only another 5 percent of the nation's food would be lost. Many environmentalists and others advocate organic farming as an alternative to heavy chemical pesticide use.

III -NONCHEMICAL CONTROLS
Many pests that are attached to crop residues can be eliminated by plowing them underground. Simple paper or plastic barriers placed around fruit trees deter insects, which can also be attracted to light traps and destroyed. Weeds can be controlled by spreading grass, leaf, or black plastic mulch. Weeds also may be pulled or hoed from the soil.

Many biological controls are also effective. Such insect pests as the European corn borer, Pyrausta nubilalis, and the Japanese beetle, Popillia japonica, have been controlled by introducing their predators and parasites. Wasps that prey on fruit-boring insect larvae are now being commercially bred and released in California orchards. The many hundreds of species of viruses, bacteria, protozoa, fungi, and nematodes that parasitize pest insects and weeds are now being investigated as selective control agents. Another area of biological control is breeding host plants to be pest resistant, making them less prone to attack by fungi and insects. The use of sex pheromones is an effective measure for luring and trapping insects.

Pheromones have been synthesized for the Mediterranean fruit fly, the melon fly, and the Oriental fruit fly. Another promising pest-control method is the release of sterilized male insects into wild pest populations, causing females to bear infertile eggs. Of these techniques, breeding host-plant resistance and using beneficial parasites and predators are the most effective. Interestingly, the combined use of biological and physical controls accounts for more pest control than chemical pesticides.

Integrated pest management (IPM) is a recently developed technology for pest control that is aimed at achieving the desired control while reducing the use of pesticides. To accomplish this, various combinations of chemical, biological, and physical controls are employed. In the past, pesticides were all too often applied routinely whether needed or not. With IPM, pest populations as well as beneficial parasite and predator populations are monitored to determine whether the pests actually present a serious problem that needs to be treated. If properly and extensively employed, IPM might reduce pesticide use by as much as 50 percent, while at the same time improving pest control. If this goal were achieved, the environmental problems would be minimized, and significant benefits would result for farmers and society as a whole

51. PROTEIN

I -INTRODUCTION

Protein, any of a large number of organic compounds that make up living organisms and are essential to their functioning. First discovered in 1838, proteins are now recognized as the predominant ingredients of cells, making up more than 50 percent of the dry weight of animals. The word protein is coined from the Greek proteios, or ―primary.‖

Protein molecules range from the long, insoluble fibers that make up connective tissue and hair to the compact, soluble globules that can pass through cell membranes and set off metabolic reactions. They are all large molecules, ranging in molecular weight from a few thousand to more than a million, and they are specific for each species and for each organ of each species. Humans have an estimated 30,000 different proteins, of which only about 2 percent have been adequately described. Proteins in the diet serve primarily to build and maintain cells, but their chemical breakdown also provides energy, yielding close to the same 4 calories per gram as do carbohydrates.

Besides their function in growth and cell maintenance, proteins are also responsible for muscle contraction. The digestive enzymes are proteins, as are insulin and most other hormones. The antibodies of the immune system are proteins, and proteins such as hemoglobin carry vital substances throughout the body.

II -NUTRITION

Whether found in humans or in single-celled bacteria, proteins are composed of units of about 20 different amino acids, which, in turn, are composed of carbon, hydrogen, oxygen, nitrogen, and sometimes sulfur. In a protein molecule these acids form peptide bonds—bonds between amino and carboxyl (COOH) groups—in long strands (polypeptide chains). The almost numberless combinations in which the acids line up, and the helical and globular shapes into which the strands coil, help to explain the great diversity of tasks that proteins perform in living matter.

To synthesize its life-essential proteins, each species needs given proportions of the 20 main amino acids. Although plants can manufacture all their amino acids from nitrogen, carbon dioxide, and other chemicals through photosynthesis, most other organisms can manufacture only some of them. The remaining ones, called essential amino acids, must be derived from food. Nine essential amino acids are needed to maintain health in humans: leucine, isoleucine, lysine, methionine, phenylalanine, threonine, tryptophan, valine, and histidine. All of these are available in proteins produced in the seeds of plants, but because plant sources are often weak in lysine and tryptophan, nutrition experts advise supplementing the diet with animal protein from meat, eggs, and milk, which contain all the essential acids.

Most diets—especially in the United States, where animal protein is eaten to excess—contain all the essential amino acids. (Kwashiorkor, a wasting disease among children in tropical Africa, is due to an amino acid deficiency.) For adults, the Recommended Dietary Allowance (RDA) for protein is 0.79 g per kg (0.36 g per lb) of body weight each day. For children and infants this RDA is doubled and tripled, respectively, because of their rapid growth.

III -STRUCTURE OF PROTEINS

The most basic level of protein structure, called the primary structure, is the linear sequence of amino acids. Different sequences of the acids along a chain, however, affect the structure of a protein molecule in different ways. Forces such as hydrogen bonds, disulfide bridges, attractions between positive and negative charges, and hydrophobic (―water-fearing‖) and hydrophilic (―waterloving‖) linkages cause a protein molecule to coil or fold into a secondary structure, examples of which are the so-called alpha helix and the beta pleated sheet. When forces cause the molecule to become even more compact, as in globular proteins, a tertiary protein structure is formed. When a protein is made up of more than one polypeptide chain, as in hemoglobin and some enzymes, it is said to have a quaternary structure.

IV -INTERACTION WITH OTHER PROTEINS

Polypeptide chains are sequenced and coiled in such a way that the hydrophobic amino acids usually face inward, giving the molecule stability, and the hydrophilic amino acids face outward, where they are free to interact with other compounds and especially other proteins. Globular proteins, in particular, can join with a specific compound such as a vitamin derivative and form a coenzyme, or join with a specific protein and form an assembly of proteins needed for cell chemistry or structure.

V -FIBROUS PROTEINS

The major fibrous proteins, described below, are collagen, keratin, fibrinogen, and muscle proteins.

A -Collagen

Collagen, which makes up bone, skin, tendons, and cartilage, is the most abundant protein found in vertebrates. The molecule usually contains three very long polypeptide chains, each with about 1000 amino acids, that twist into a regularly repeating triple helix and give tendons and skin their great tensile strength. When long collagen fibrils are denatured by boiling, their chains are shortened to form gelatin.

B -Keratin
Keratin, which makes up the outermost layer of skin and the hair, scales, hooves, nails, and feathers of animals, twists into a regularly repeating coil called an alpha helix. Serving to protect the body against the environment, keratin is completely insoluble in water. Its many disulfide bonds make it an extremely stable protein, able to resist the action of proteolytic (protein-hydrolyzing) enzymes. In beauty treatments, human hair is set under a reducing agent, such as thioglycol, to reduce the number of disulfide bonds, which are then restored when the hair is exposed to oxygen.

C -Fibrinoge
Fibrinogen is a blood plasma protein responsible for blood clotting. With the catalytic action of thrombin, fibrinogen is converted into molecules of the insoluble protein fibrin, which link together to form clots.

D -Muscle Proteins
Myosin, the protein chiefly responsible for muscle contraction, combines with actin, another muscle protein, forming actomyosin, the different filaments of which shorten, causing the contracting action.

VI -GLOBULAR PROTEINS
Unlike fibrous proteins, globular proteins are spherical and highly soluble. They play a dynamic role in body metabolism. Examples are albumin, globulin, casein, hemoglobin, all of the enzymes, and protein hormones. The albumins and globulins are classes of soluble proteins abundant in animal cells, blood serum, milk, and eggs. Hemoglobin is a respiratory protein that carries oxygen throughout the body and is responsible for the bright red color of red blood cells. More than 100 different human hemoglobins have been discovered, among which is hemoglobin S, the cause of sickle-cell anemia, a hereditary disease suffered mainly by blacks.

A -Enzymes
All of the enzymes are globular proteins that combine rapidly with other substances, called substrate, to catalyze the numerous chemical reactions in the body. Chiefly responsible for metabolism and its regulation, these molecules have catalytic sites on which substrate fits in a lock-and-key manner to trigger and control metabolism throughout the body.

B -Protein Hormones
These proteins, which come from the endocrine glands, do not act as enzymes. Instead they stimulate target organs that in turn initiate and control important activities—for example, the rate of metabolism and the production of digestive enzymes and milk. Insulin, secreted by the islands of Langerhans, regulates carbohydrate metabolism by controlling blood glucose levels. Thyroglobulin, from the thyroid gland, regulates overall metabolism; calcitonin, also from the thyroid, lowers blood calcium levels. Angiogenin, a protein structurally determined in the mid-1980s, directly induces the growth of blood vessels in tissues.

C -Antibodies
Also called immunoglobulins, antibodies (see Antibody) make up the thousands of different proteins that are generated in the blood serum in reaction to antigens (body-invading substances or organisms). A single antigen may elicit the production of many antibodies, which combine with different sites on the antigen molecule, neutralize it, and cause it to precipitate from the blood.

D -Microtubules
Globular proteins can also assemble into minute, hollow tubes that serve both to structure cells and to conduct substances from one part of a cell to another. Each of these microtubules, as they are called, is made up of two types of nearly spherical protein molecules that pair and join onto the growing end of the microtubule, adding on length as required. Microtubules also make up the inner structure of cilia, the hairlike appendages by which some microorganisms propel themselves.

52. VERTEBRATES

I -INTRODUCTION
Vertebrate, animal with a backbone, or spinal column, made of interlocking units called vertebrae. This strong but flexible structure supports the body and anchors the limbs, and it also protects the nerves of the spinal cord. Vertebrates include fish, amphibians, and reptiles, as well as birds and mammals. In all vertebrates, the spinal column forms part of a complete internal skeleton. Unlike the hard external skeleton covering an insect, which is periodically shed as the insect grows, a vertebrate's internal skeleton can grow gradually along with the rest of the body.

Vertebrates make up only about 2 percent of the animal species, and they belong to just 1 of more than 30 phyla, or overall groups, in the animal kingdom. Despite this, vertebrates occupy a dominant position in almost all habitats and are by far the most familiar animals. When asked to name an animal at random, most people identify a type of vertebrate.

There are several reasons why vertebrates are so successful and so noticeable. One has to do with their size. Invertebrates—that is, animals without backbones, such as worms, shellfish, and insects—tend to be small and slow moving. This is because they lack effective ways to support a large body and the muscles needed to power it. Vertebrates, on the other hand, have evolved a much more versatile support system. Their skeletons can be adapted for use in many different ways and work just as well in an animal weighing 4 tons as in one weighing 113 g (4 oz). As a result, vertebrates have been able to develop bigger, faster bodies than invertebrates.

Vertebrates also have highly developed nervous systems. With the help of specialized nerve fibers, they can react very quickly to changes in their surroundings, giving them a competitive edge.

II -CHARACTERISTICS

In nearly all vertebrates, bone gives the skeleton its strength. Bone is a living tissue composed of hard mineral salts produced by specialized cells. Unlike an oyster's shell or a grasshopper's body case, bone can strengthen after it has reached full size, and it can be repaired if it breaks. The only vertebrates that do not have this kind of skeleton are cartilaginous fish, a group that includes sharks, skates, and rays. As their name suggests, the skeletons of these species are made of cartilage, a rubbery tissue that other vertebrates have mainly in their joints.

A vertebrate's spinal column is held together by strong ligaments, but the faces of adjoining vertebrae are separated by elastic pads called intervertebral disks. These disks allow a small amount of movement at each joint, and as a result the entire spine can bend. How far the spine bends depends on the number of vertebrae that compose it and how they are shaped. Frogs, for example, can have as few as nine vertebrae, and their backbones hardly bend at all. Humans have 33 vertebrae, making us fairly flexible, and some snakes have more than 400, enabling them to shape their bodies into coils.

Besides the backbone, vertebrates share many other physical features. Their bodies are more or less bilaterally symmetrical (divisible into two equal halves), with sense organs concentrated in the head. Most vertebrates have jaws, and their brains are usually protected by a bony case called the cranium. Most also have limbs, but the shapes and uses of vertebrate limbs vary enormously. Fish typically have several paired fins and a large finned tail, but all other vertebrates either have four limbs or are descended from ancestors that had four. Four-limbed animals, known as tetra pods, use their limbs to swim, walk, run, and fly.

Although vertebrates do not have external skeletons, they often have other anatomical features that protect the surface of their bodies. Most fish and reptiles have a covering of hard scales, while birds and mammals have feathers or hair. Feathers and hair are not as tough as scales, but they have other functions apart from physical protection. One of the most important is insulation. By regulating the heat generated inside the body, such coverings allow birds and mammals to remain active in a wide range of temperatures.

Nearly all vertebrates breed by sexual reproduction, either laying eggs or giving birth to live young. The few exceptions to this rule include animals such as North American whiptail lizards, which can breed without mating in a process known as parthenogenesis. In several species of these lizards, males have never been found.

III -TYPES OF VERTEBRATES

There are over 40,000 species of vertebrates, which scientists classify into five groups: (1) fish, (2) amphibians, (3) reptiles, (4) birds, and (5) mammals. Scientists divide fish into three groups based on their anatomy: jawless fish, cartilaginous fish, and bony fish. The other vertebrate groups are made up of tetrapods, which have lungs and generally live on land.

A -Jawless Fish

Jawless fish are the only living vertebrates that have never evolved jaws. There are about 50 species—a tiny fraction of the world's total fish—and they are instantly recognizable by their suckerlike mouths. Eels, lampreys, and hagfish are examples of jawless fish.

B -Cartilaginous Fish

Cartilaginous fish do have jaws and use them to deadly effect. Numbering about 1,000 species, they include sharks, skates, and rays, as well as chimaeras, also known as ratfish. Cartilaginous fish are widespread throughout the world's oceans. Most skates and rays feed on or near the seabed, but sharks typically hunt in open water.

C -Bony Fish

Bony fish are some of the most successful vertebrates alive today. These animals can be found in a vast variety of habitats, from coral reefs and the deep-sea bed to lakes hidden away in caves. As their name indicates, bony fish have a skeleton made of bone, and most also have an air-filled sac called a swim bladder that keeps them buoyant. At least 24,000 species of bony fish have been identified, and many more probably await discovery. Common bony fish include salmon, sturgeon, and cod.

D -Amphibians

Amphibians make up the smallest of the four groups of tetrapods, with about 4,000 species. Most amphibians, such as frogs and toads, live in damp habitats. Like fish, the majority of amphibians reproduce by laying eggs. Amphibians usually lay their eggs in water, because they dry out quickly in air. The eggs produce swimming, fishlike young called tadpoles, which develop limbs and lungs as they mature.

E -Reptiles

Compared to amphibians, reptiles are much more fully adapted to life on land. They have scaly, waterproof skin, and they either give birth to live young or lay eggs with waterproof shells. There are about 7,000 species alive today, including snakes, alligators, and turtles. During the age of the dinosaurs, about 230 million to 65 million years ago, reptiles outnumbered all other land vertebrates put together.

F -Birds

Birds evolved from flightless reptiles but underwent some major changes in body form during their evolution. Of the roughly 10,000 species alive today, most have lightweight, air-filled bones, and all have a unique and highly efficient respiratory system that is found in no other group of vertebrates.

G -Mammals
Mammals are the only vertebrates that raise their young by feeding them on milk produced by the mother's body, and the only ones that have teeth that are individually specialized for particular functions. Mammal species number about 4,600, and they include the largest animals on land and in the sea. Dogs, bears, monkeys, whales, and humans are all mammals.

IV -THE ORIGIN OF VERTEBRATES
Biologists believe that vertebrates evolved over millions of years from animals similar to today's lancelets, which burrow in sand on the seabed and filter food from the water. Lancelets possess certain traits similar to vertebrates, including a reinforcing structure called a notochord that runs the length of the body. In a lancelet the notochord is the only hard part of the body, and it allows the animal to wriggle without losing its shape. In most vertebrates, the notochord is lost during early development, and its role is taken over by bone. The characteristics shared by lancelets and vertebrates cause scientists to classify them together in the chordate phylum.

Scientists do not know exactly how the transition from lancelet to vertebrate occurred. Fossils of fishlike animals found in China indicate that vertebrates evolved at the start of the Cambrian Period, an interval of geologic time that began about 570 million years ago. These fish lacked a bony skeleton and teeth (scientists propose that their skeletal structures were made of cartilage), but they did have gill slits and a muscle arrangement similar to today's fish. Once vertebrates evolved hard body parts, they began to leave more fossilized remains. Fish called ostracoderms, which had bony plates covering their bodies, first appeared in the late Cambrian Period, about 500 million years ago. Like present-day lampreys and hagfish, ostracoderms had no jaws. They probably fed by sucking water into their mouths and then swallowing any food it contained.

With the evolution of jaws, vertebrates acquired a valuable new asset in the struggle for survival, one that enabled them to collect food in a variety of different ways. Jaws first appeared in fish about 420 million years ago, during the mid-Silurian Period. Unlike earlier vertebrates, jawed fish developed complex internal skeletons and paired fins, which helped them maneuver as they pursued their food or escaped from their enemies. Over time, evolution has produced vertebrates with many different body types and behaviors. As a result, vertebrates can now be found in almost every part of the world

53. INVERTEBRATES

I -INTRODUCTION
Invertebrate, any animal lacking a backbone. Invertebrates are by far the most numerous animals on Earth. Nearly 2 million species have been identified to date. These 2 million species make up about 98 percent of all the animals identified in the entire animal kingdom. Some scientists believe that the true number of invertebrate species may be as high as 100 million and that the work of identifying and classifying invertebrate life has only just begun.

Invertebrates live in a vast range of habitats, from forests and deserts to caves and seabed mud. In oceans and lakes they form part of the plankton—an immense array of miniature living organisms that drift in the surface currents. Invertebrates are also found in the soil beneath our feet and in the air above our heads. Some are powerful fliers, using wings to propel themselves, but others, particularly the smallest invertebrates, float on the slightest breeze. These tiny invertebrates form clouds of aerial plankton that drift unseen through the skies.

Although the majority of invertebrates are small, a few reach impressive sizes. The true heavyweights of the invertebrate world are giant squid, which can be over 18 m (60 ft) long and can weigh more than 2,000 kg (4,000 lb). The longest are ribbon worms, also known as nemerteans, whose pencil-thin bodies can grow up to 55 m (180 ft) from head to tail. At the other end of the size scale, animals called rotifers rank among the smallest invertebrates of all. Some species may reach 3 mm (0.12 in) in size, but most are less than 0.001 mm (0.00004 in), smaller than the largest bacteria.

II -PHYSICAL CHARACTERISTICS
Due to their numbers and variety, invertebrates share only a single trait in common: the absence of a backbone. Many invertebrates have no hard body parts at all. These soft-bodied invertebrates, which include earthworms, keep their shape by maintaining an internal pressure, similar to the air pressure within an inflated balloon. However, having a soft body has disadvantages, one of which is that it leaves animals vulnerable to attack from predators.

To defend against predators, other invertebrates have evolved exoskeletons, hard outer coverings such as the shells found in clams and mussels and the body cases that surround adult insects. As well as protecting the animal, these exoskeletons also provide anchorage for muscles. On land, a body case is also useful because it prevents the water that bathes internal structures from evaporating. As a result the animal does not dry up and die. Arthropods, animals with a hard, outer skeleton and a jointed body and limbs, make up the single largest group of invertebrates. Arthropods include insects, crustaceans, and arachnids, such as spiders and ticks.

Invertebrates have two basic body plans. Some invertebrates, such as corals and sea anemones, have a circular body plan arranged around a central mouth, similar to the way spokes radiate out from the hub of a wheel. This type of body plan is known as radial symmetry. Animals with radial symmetry often spend their adult lives fastened in one place, like the sea anemone that attaches to a rock, waiting for food to pass by. By contrast, invertebrates that move in search of food, such as flatworms, have an elongated body plan known as bilateral symmetry. Invertebrates with bilateral symmetry have right and left halves that mirror each other, and they typically have a definite front and back end. They have a head that often contains one or more pairs of eyes, together with organs that can taste, smell, or touch. However, major sense organs are often found on other body parts among some invertebrates. Katydids, for example, have hearing organs on their front legs, just below knee like joints.

Compared to vertebrates (animals with backbones), most invertebrates have simple nervous systems, and they behave almost entirely by instinct. This system works well most of the time, even though these animals cannot learn from their mistakes. Moths, for example, repeatedly flutter around bright lights, even at the risk of getting burned. Notable exceptions are octopuses and their close relatives, which are thought to be the most intelligent animals in the invertebrate world. Studies have shown that these animals have the ability to learn. In some experiments they have solved simple puzzles, such as opening containers to retrieve food.

Invertebrates differ from each other internally in a wide variety of ways. Some have respiratory organs, circulatory systems, and excretory organs for getting rid of waste. The simplest invertebrates, such as placozoans, survive with few or no specialized organs at all. These animals absorb what they need from their surroundings—a way of life that works only in watery habitats and only with small animals.

III -TYPES OF INVERTEBRATES

Zoologists (scientists who study animals) classify invertebrates into about 30 major groups, known as phyla. These phyla vary enormously in the number of species they contain. Arthropods (phylum Arthropoda) are the invertebrate phylum with the most species—more than one million known species and countless more awaiting discovery. The mollusks (phylum Mollusca) make up the second largest group of invertebrates, with at least 50,000 species. Among the simplest invertebrates are the sponges (phylum Porifera). Other major invertebrate phyla include the cnidarians (phylum Cnidaria), echinoderms (phylum Echinodermata), and several different groups of worms, including flatworms (phylum Platyhelminthes), roundworms (phylum Nematoda), and annelids (phylum Annelida).

Arthropods live in every habitat on Earth from mountaintops to hydrothermal vents, springs of hot water located on the deep ocean floor. Surrounded by protective exoskeletons, arthropods have tubular legs that bend at flexible joints. This unique characteristic sets them apart from all other invertebrates, and it enables them to hop, walk, and run.

Insects dominate the arthropod phylum. Making up 90 percent of all arthropods, insects have a strong claim to be the most successful animals in the world. On land, they live in almost every habitat, aided by their small size and, for many, their ability to fly. They also live in fresh water, but remarkably, they have failed to colonize the sea. Some zoologists believe this is because crustaceans have already exploited this habitat to its fullest.

Mollusks make up the second largest group of invertebrates. Even by invertebrate standards mollusks are extremely varied. Mollusks include snails, clams, octopuses, and squid, as well as some lesser-known animals, such as chitons and monoplacophorans. Some mollusks, such as bivalves, are sedentary animals, while others such as squid are jet-propelled predators that are the swiftest swimmers in the invertebrate world. Most sedentary mollusks are filter feeders—that is, they feed on tiny organisms that they strain from water. Other mollusks, including snails and other gastropods, scrape up their food using a radula—a ribbonlike mouthpart that is unique to mollusks and covered with rows of microscopic teeth.

Sponges have many unique characteristics that set them apart from other kinds of animal life. They are the only animals with skeletons made of microscopic mineral spikes and the only ones that feed by pumping water through hollow pores. Some of their cells are remarkably like free-living protozoans called collar flagellates. To evolutionary biologists, this resemblance strongly suggests that sponges and other invertebrates arose from protozoan-like ancestors.

Cnidarians include jellyfish, sea anemones, and corals. Their bodies have two layers of cells, a central digestive cavity, and a mouth surrounded by stinging tentacles. Most cnidarians are quite small, but the largest jellyfish—a species from the North Atlantic Ocean—can grow over 2 m (7 ft) across, with tentacles over 30 m (100 ft) long.

Among the major phyla, the echinoderms are the most distinctive and unusually shaped. They include starfish, sea urchins, and sea cucumbers and are the only animals with a five-pointed design. They live in the sea and move with the help of tiny fluid-filled feet—another feature found nowhere else in the animal world.

Zoologists recognize several different groups of worms. The phylum known as flatworms contains the simplest animals possessing heads. Nerves and sense organs are concentrated in the head. Most flatworms are paper-thin and live in a variety of wet or damp habitats, including the digestive systems of other animals. Roundworms represent another phylum. They are more complex than flatworms, with cylindrical bodies and mouthparts designed to pierce their food. Although flatworms have digestive systems with only one opening, the roundworm digestive system runs from the mouth straight through its body to an excretory opening—a body plan shared by more advanced invertebrates as well as vertebrates.

Although roundworms are extremely abundant, they often go unseen. So, too, do many worms that live exclusively in the sea, such as spoonworms (phylum Echiura), peanut worms (phylum Sipuncula), and pogonophores (phylum Pogonophora). Annelids are a large group of worms that contain some more familiar species. Among them are earthworms—annelids that feed by burrowing through the soil. An earthworm's body is divided into repeated segments or rings, a feature shared by annelids as a whole.

IV -REPRODUCTION AND LIFE CYCLE

Invertebrates display a wide variety of methods of reproduction. Some invertebrates reproduce by asexual reproduction, in which all offspring are genetically identical to the parent. Asexual reproduction methods include fragmentation, in which animals divide into two or more offspring, and budding, in which animals sprout buds that break away to take up life on their own. The majority of invertebrates reproduce sexually. The genes from two parents recombine to produce genetically unique individuals. For most invertebrates, sexual reproduction involves laying eggs. With a few exceptions, such as scorpions and spiders, most invertebrates abandon their eggs as soon as they are laid, leaving them to develop on their own.

When invertebrate eggs hatch, the animals that emerge often look nothing like their parents. Some are so different that, in the past, zoologists mistook them for entirely new species. Young like this are known as larvae. As they grow up, larvae change shape, a process known as metamorphosis. A larval stage enables invertebrates to live in different habitats at different stages of their lives. For example, adult mussels live fastened to rocks, but their larvae live floating among plankton. By having larvae that drift with the currents, mussels are able to disperse and find homes with new food sources for their adult life.

The change from larva to adult is quite gradual in many invertebrates, such as crabs and lobsters, but in insects it can be much more abrupt. Caterpillars, the larvae of butterflies and moths, often live for several months, but they take just a few days to turn into adults. During the transition stage, known as the pupa, the caterpillar's body is broken down and reassembled, forming an adult insect that is ready to breed.

Most invertebrates are short-lived animals, but slow-growing species often break this rule. Wood-boring beetles can live well into their teens, while queen termites can live 40 years or more. But in the invertebrate world, the real veterans live in the sea. Growth lines on bivalve shells suggest that some clams can live to be 400 years old or more. An age of about 200 years has been claimed for pogonophoran worms living around hydrothermal vents in the darkness of the deep seafloor.

V -EVOLUTION

As the simplest animals, invertebrates date back to the time when animal life first began in ancient shallow seas. Zoologists are uncertain when this was, because the first invertebrates were small and soft-bodied and left no direct fossil remains. However, some scientists believe that strange patterns preserved in sedimentary rocks dating back to 1 billion years ago may be the fossilized tracks and burrows of ancient invertebrates. Other scientists, studying genetic material in living animals, believe that the earliest invertebrates may have appeared even earlier and may already have begun to separate into different phyla before 1 billion years ago.

The oldest recognized fossils of invertebrates date back to the close of the Precambrian period, about 550 million years ago. The best known of this fossil finds, from the Ediacaran Hills in southern Australia, include animals that look like jellyfish and annelid worms. Zoologists disagree about their status. Some think that they might well be ancestors of animals alive today, but others believe they belong to a group of invertebrates that eventually became extinct.

With the start of the Cambrian period 542 million years ago, invertebrate life evolved with almost explosive speed. Due to the appearance of the first invertebrates with exoskeletons, the fossil record provides a rich record of invertebrate life in the Cambrian period. By the time the Cambrian period ended 488 million years ago, all the invertebrate phyla alive today were established.

Between that time and the present, invertebrates spread through the seas and also invaded land. Scientists believe that the first land dwellers were almost certainly arthropods, including the forerunners of wingless insects. During the Carboniferous period, which began 359 million years ago, flying insects appeared, including giant dragonflies with a wingspan of up to 75 cm (30 in). But on land the great expansion of invertebrate life occurred during the Cretaceous period, which started 145 million years ago. Flowering plants first evolved in this period, enabling insects to exploit a whole new source of food and triggering a huge growth in insect life that has continued to this day.

While many invertebrates flourished, some of the most successful groups of invertebrates in the fossil record nonetheless became extinct. Giant sea scorpions and trilobites were types of arthropods that thrived for much of the Paleozoic era, about 270 million years ago, but were unable to survive the great mass extinction at the end of the Permian period 251 million years ago. Ammonites (mollusks related to today's octopuses and squids) fared better. They first appeared during the Silurian period about 440 million years ago and lived into the Mesozoic era, only to vanish at the same time as the dinosaurs, about 65 million years ago. Their intricate massive spiral shells were often superbly preserved as fossils, some measuring almost 2 m (7 ft) across.

VI -IMPORTANCE OF INVERTEBRATES

The continued prominence of invertebrates, measured by their great diversity and abundance, indicates that these animals have adapted to their ecosystems over millions of years. In so doing, invertebrates have become necessary to the health of Earth's ecology. For instance, all ecosystems support one or more food chains that form food webs. Each chain begins with plants, known as primary producers, which convert light energy into food. Primary producers are eaten by primary consumers, and secondary consumers eat the plant-eating primary consumers. Decomposers derive their energy from the dead remains of plants and animals. Invertebrates occupy several niches in this food web, acting as primary consumers, secondary consumers, and decomposers.

Many invertebrates have a direct and invaluable impact on their environment. For example, the common earthworm burrows deep below the surface, consuming soil along the way. Coiled soil masses known as casts are excreted from the worm's digestive system, making the soil more fertile. The earthworm's burrowing action continually moves mineral-rich soil to the surface, which improves plant growth. The burrowing action also aerates soil, enhancing drainage. In another example, as honey bees, butterflies, and moths flit from flower to flower collecting nectar, they inadvertently transport pollen from the male reproductive structure of one flower to the female reproductive structure of another flower. Known as pollination, this leads to the fertilization of the plant's seeds—an essential stage in the process of reproduction.

Other invertebrates form mutually beneficial partnerships with other animals. For example, some crabs form alliances with sea anemones, which they fasten to their backs. In this alliance, the crab is protected from predators by the anemone's stinging tentacles. The anemone, in turn, receives food particles as the crab tears up meat from the animals it consumes. As the crab grows, it periodically sheds its body case. Before doing so, it removes the anemone, and then afterwards puts it back in place.

Humans sometimes share a troubled relationship with invertebrates. A number of invertebrate organisms cause many parasitic diseases in humans and farm animals. These parasites survive by feeding and reproducing inside a host, often causing internal destruction. Some of the most damaging parasites include the flatworm Schistosoma, which causes schistosomiasis; the roundworms that cause hookworm infection; and the roundworm larvae of Trichinella spiralis that cause trichinosis. Other invertebrates are agricultural pests, destroying plant crops. Insects such as leaf beetles, flea beetles, and young caterpillars feed on the leaves, stems, roots, and flowers of plants. Sucking insects, including aphids, leafhoppers, and scales, remove plant sap, weakening the plants. Sucking insects can also spread disease-causing viruses and bacteria to plants. The larvae and adult stages of some roundworms are parasites of plants. Using specialized structures called stylets, these roundworms pierce plants at the roots to extract cell content, killing the plant.

Although invertebrates can cause problems for humans, they are more often beneficial. In many cultures, invertebrates such as squid, octopuses, cuttlefish, clams, mussels, crabs, and lobsters are considered popular food items. Scientists use invertebrates for a variety of experiments that have profound benefits for human health. Scientists have performed delicate surgery on the glandular systems of caterpillars and roaches to uncover clues to the function of glands in humans. In other experiments, scientists have given spiders different types of drugs and observed the animals as they created spider webs. The different pattern of spider webs offered a way to distinguish and measure the effects of various drugs.

The vinegar fly Drosophila melanogaster, also known as the fruit fly, has long been the standard test subject in the field of genetics. In the 1910s and 1920s American geneticist Thomas Hunt Morgan used the vinegar fly to demonstrate that genes lie in a linear fashion on chromosomes, establishing the chromosomal basis of inheritance. In early 2000 studies of vinegar flies continued to advance the field of modern genetics when researchers sequenced the fly's entire genetic makeup, or genome. The techniques used to reveal the vinegar fly genome were then applied to the efforts to decode the human genome

54. LIVER

I -INTRODUCTION
Liver, largest internal organ of the human body. The liver, which is part of the digestive system, performs more than 500 different functions, all of which are essential to life. Its essential functions include helping the body to digest fats, storing reserves of nutrients, filtering poisons and wastes from the blood, synthesizing a variety of proteins, and regulating the levels of many chemicals found in the bloodstream. The liver is unique among the body's vital organs in that it can regenerate, or grow back, cells that have been destroyed by some short-term injury or disease. But if the liver is damaged repeatedly over a long period of time, it may undergo irreversible changes that permanently interfere with function.

II -STRUCTURE OF THE LIVER
The human liver is a dark red-brown organ with a soft, spongy texture. It is located at the top of the abdomen, on the right side of the body just below the diaphragm—a sheet of muscle tissue that separates the lungs from the abdominal organs. The lower part of the rib cage covers the liver, protecting it from injury. In a healthy adult, the liver weighs about 1.5 kg (3 lb) and is about 15 cm (6 in) thick.

Despite its many complex functions, the liver is relatively simple in structure. It consists of two main lobes, left and right, which overlap slightly. The right lobe has two smaller lobes attached to it, called the quadrate and caudate lobes.

Each lobe contains many thousands of units called lobules that are the building blocks of the liver. Lobules are six-sided structures each about 1 mm (0.04 in) across. A tiny vein runs through the center of each lobule and eventually drains into the hepatic vein, which carries blood out of the liver. Hundreds of cubed-shaped liver cells, called hepatocytes, are arranged around the lobule's central vein in a radiating pattern. On the outside surface of each lobule are small veins, ducts, and arteries that carry fluids to and from the lobules. As the liver does its work, nutrients are collected, wastes are removed, and chemical substances are released into the body through these vessels.

Unlike most organs, which have a single blood supply, the liver receives blood from two sources. The hepatic artery delivers oxygen-rich blood from the heart, supplying about 25 percent of the liver's blood. The liver also receives oxygen-depleted blood from the hepatic portal vein. This vein, which is the source of 75 percent of the liver's blood supply, carries blood to the liver that has traveled from the digestive tract, where it collects nutrients as food is digested. These nutrients are delivered to the liver for further processing or storage.

Tiny blood vessel branches of the hepatic artery and the hepatic portal vein are found around each liver lobule. This network of blood vessels is responsible for the vast amount of blood that flows through the liver—about 1.4 liters (about 3 pt) every minute. Blood exits the liver through the hepatic vein, which eventually drains into the heart.

III -FUNCTIONS OF THE LIVER
One of the liver's primary jobs is to store energy in the form of glycogen, which is made from a type of sugar called glucose. The liver removes glucose from the blood when blood glucose levels are high. Through a process called glycogenesis, the liver combines the glucose molecules in long chains to create glycogen, a carbohydrate that provides a stored form of energy. When the amount of glucose in the blood falls below the level required to meet the body's needs, the liver reverses this reaction, transforming glycogen into glucose.

Another crucial function of the liver is the production of bile, a yellowish-brown liquid containing salts necessary for the digestion of lipids, or fats. These salts are produced within the lobules. Bile leaves the liver through a network of ducts and is transported to the gallbladder, which concentrates the bile and releases it into the small intestine. Vitamins are also stored in the liver.
Drawing on the nutrient-rich blood in the hepatic portal vein, the liver collects and stores supplies of vitamins A, D, E, and K. The B vitamins are also stored here, including a two- to four-year supply of Vitamin B12.

The liver also functions as the body's chemical factory. Several important proteins found in the blood are produced in the liver. One of these proteins, albumin, helps retain calcium and other important substances in the bloodstream. Albumin also helps regulate the movement of water from the bloodstream into the body's tissues. The liver also produces globin, one of the two components that form hemoglobin—the oxygen-carrying substance in red blood cells. Certain globulins, a group of proteins that includes antibodies, are produced in the liver, as are the proteins that make up the complement system, a part of the immune system that combines with antibodies to fight invading microorganisms.

Many other chemicals are produced by the liver. These include fibrinogen and prothrombin, which help wounds to heal by enabling blood to form clots, and cholesterol, a key component of cell membranes that transports fats in the bloodstream to body tissues. In addition to manufacturing chemicals, the liver helps clear toxic substances, such as drugs and alcohol, from the bloodstream. It does this by absorbing the harmful substances, chemically altering them, and then excreting them in the bile.

IV -LIVER DISEASES

Although the liver is exposed to many potentially harmful substances, it is a remarkable organ that is able to regenerate, or repair or replace, injured tissue. Its construction, in which many lobules perform the same task, means that if one section of the liver is damaged, another section will perform the functions of the injured area indefinitely or until the damaged section is repaired. But the liver is subject to many diseases that can overwhelm its regeneration abilities, threatening a person's health.

Diseases of the liver range from mild infection to life-threatening liver failure. For many of these ailments, the first sign of a problem is a condition called jaundice, characterized by a yellowish coloring of the skin and the whites of the eye. It develops when liver cells lose their ability to process bilirubin, the yellowish-brown pigment found in bile.

The liver can be harmed whenever injury or disease affects the rest of the body. For example, cancer may spread from the stomach or intestines to the liver, and diabetes, if not properly treated, may result in damage to the liver. Some diseases caused by parasites, including amebiasis and schistosomiasis, can damage the liver. Drug use, including long-term use of some prescription medications as well as illegal drugs, can also cause liver damage. Poisons can easily damage liver cells and even cause complete liver failure, especially the poisons found in certain mushrooms.

One of the most common liver diseases is hepatitis, an inflammation of the liver. Hepatitis may be caused by exposure to certain chemicals, by autoimmune diseases, or by bacterial infections. But hepatitis is most often caused by one of several viruses. The hepatitis A virus (HAV) can produce flu like symptoms and jaundice, but many people who contract it have no symptoms. The disease tends to resolve on its own. Because HAV lives in feces in the intestinal tract, hepatitis A is prevalent in areas where drinking water is contaminated with raw sewage. Good hygiene practices and a hepatitis A vaccination are effective measures of prevention. Hepatitis B is a more serious ailment. Unlike HAV, hepatitis B virus (HBV) may remain active in the body for many years after the time of infection, sometimes permanently damaging the liver. HBV is found in blood and other body fluids—tears, saliva, and semen—and is spread through unprotected sexual intercourse and the sharing of infected needles or other sharp objects that puncture the skin.

In developed countries, alcohol-induced liver diseases far outnumber hepatitis and all other liver disorders. Heavy alcohol use causes fat deposits to build up in the liver, possibly leading to chronic hepatitis, which causes scarring and destruction of liver cells. Over many years, scarring in the liver can progress to cirrhosis, a disease characterized by diminished blood flow through this important organ. When this occurs, toxins are not adequately removed from the blood, blood pressure increases in the hepatic portal vein, and substances produced by the liver, such as blood proteins, are not adequately regulated. Cirrhosis cannot be reversed, but liver function can significantly improve in people who stop consuming alcohol during the early stages of this condition. Beyond abstinence from alcohol, treatments for cirrhosis may include drug therapy or surgery to redirect blood flow.

For people with severe liver disease or impending liver failure, organ transplantation may be an option. Unlike some organ transplants, such as kidney transplants, liver transplants are complex procedures that have not had high long-term success rates. Fortunately, new techniques and drugs are improving the outcome of liver transplants. Current success rates range between 60 and 80 percent, with more than half of recent transplant recipients surviving more than five years. Most of these people have an excellent prognosis for leading healthy, normal lives.

Liver Picture

The largest internal organ in humans, the liver is also one of the most important. It has many functions, among them the synthesis of proteins, immune and clotting factors, and oxygen and fat-carrying substances. Its chief digestive function is the secretion of bile, a solution critical to fat emulsion and absorption. The liver also removes excess glucose from circulation and stores it until it is needed. It converts excess amino acids into useful forms and filters drugs and poisons from the bloodstream, neutralizing them and excreting them in bile. The liver has two main lobes, located just under the diaphragm on the right side of the body. It can lose 75 percent of its tissue (to disease or surgery) without ceasing to function

Healthy and Diseased Livers

The liver cells on the left are from a healthy liver, while the cells on the right came from the liver of a person with cirrhosis of the liver. Cirrhosis is usually caused by toxins (including alcohol) in the blood or by hepatitis. In cirrhosis, dead and damaged liver cells are replaced by fibrous tissue, which can form masses of scar tissue and dramatically change the structure of the liver.
These fibrous areas can slow the flow of blood through the liver

55. ENZYMES

Protein molecules are built up by enzymes which join together tens or hundreds of amino acid molecules. These proteins are added to the cell membrane, to the cytoplasm or to the nucleus of the cell. They may also become the proteins which act as enzymes.

Enzymes are proteins in nature that act as catalysts. They are made in all living cells. A catalyst is a chemical substance which speeds up a reaction but does not get used up during the reaction, thus, one enzyme can be used many times over. Without these catalysts, which speed the rate of chemical reactions, metabolism would not occur at a fast enough rate to sustain life. For instance, if starch is mixed with water it will break down very slowly to sugar, taking several years. In your saliva, there is an enzyme called amylase which can break down starch to sugar in minutes or seconds.

Reactions in which large molecules are built up from smaller molecules are called anabolic reactions, whereas, reactions which split large molecules into smaller ones are called catabolic reactions.

Enzymes are specific

This means simply that an enzyme which normally acts on one substance will not act on a different one. The shape of an enzyme decides what substances it combines with. Each enzyme has a shape which exactly fits the substances on which it acts, but will not fit (or react with) the substances of different shapes.

An enzyme molecule has a dent in it called the active site. This active site is exactly the right size and shape for a molecule of the substrate to fit into (exactly like lock and key). Thus, an enzyme which breaks down starch to maltose will not also break down proteins to amino acids. Also, if a reaction takes places in stages, e.g.

starch → maltose (stage 1)

maltose → glucose (stage 2)

a different enzyme is needed for each stage.

The names of enzymes usually end with -ase and they are named according to the substance on which they act, or the reaction which they promote. For example, an enzyme which acts on proteins may be called a protease; one which removes hydrogen from a substance is a dehydrogenate.

The substance on which an enzyme acts is called its substrate. Thus, the enzyme sucrose acts on the substrate sucrose to produce the monosaccharide glucose and fructose.

Enzymes and temperature

A rise in temperature increases the rate of most chemical reactions; a fall in temperature slows them down. In many cases a rise of 10 degree Celsius will double the rate of reaction in a cell. This is equally true for enzymes controlled reactions. Between 0-50 degree Celsius, increasing the temperature increases the rate of reaction. This is because the enzyme molecules and substrate molecules move faster at higher temperatures, colliding into each other more often. But above 50 degree Celsius the enzymes, being proteins, are denatured (i.e. the shape of enzymes are changed and the enzymes can no longer combine with the substances or fit into the active site) and stop working. A denatured enzyme cannot act as a catalyst.

This is one of the reasons why organisms may be killed by prolonged exposure to high temperatures. The enzymes in their cells are denatured and the chemical reactions proceed too slowly to maintain life.

One way to test whether a substance is an enzyme is to heat it to the boiling point. If it can still carry out its reactions after this, it cannot be an enzyme. This technique is used as a _control‟ in enzyme experiment.

Enzymes and pH

pH is a measure of how acidic or alkaline a solution is. The scale runs from 1 to 14. A pH of 7 is neutral. A pH below 7 is acidic and a pH above 7 is alkaline.

Acid or alkaline conditions alter the chemical properties of proteins, including enzymes. For most enzymes, there is a small range of pH which their molecules are exactly the right shape to catalyse their reaction. Above or below this pH, their molecules lose their shape, so the substance can not fit into the enzyme‟s active site and cannot act as a catalyst. The protein digesting enzyme in your stomach, for example, works well at an acidity of pH 2. At this pH, the enzyme amylase, from your saliva, cannot work at all. Inside the cells, most enzymes will work best in neutral conditions (pH 7).

Although changes in pH affect the activity of enzymes, these effects are usually reversible, i.e. an enzyme which is inactivated by a low pH will resume its normal activity when its optimum pH is restored. Extremes of pH, however, may denature some enzymes irreversibly.

N.B. An enzyme which is denatured by extreme pH or temperature values will not resume its normal activity by decreasing pH or temperature values, as at higher temperatures and certain pH values they lose their shape and die.

The pH or temperature at which an enzyme works best is often called its optimum pH or temperature.

Rates of enzyme reactions

As explained above, the rate of an enzyme-controlled reaction depends on the temperature and pH. It also depends on the concentrations of the enzyme and its substrate. The more enzyme molecules produced by a cell, the faster the reaction will proceed, provided there are enough substrate molecules available. Similarly, an increase in the substrate concentration will speed up the reaction if there are enough enzyme molecules to cope with the additional substrate.

Intra- and extracellular enzymes

All enzymes are made inside cells. Most of them remain inside the cell to speed up the reactions in the cytoplasm and nucleus. These are called intracellular enzymes (_intra‟ means _inside‟). In a few cases, the enzymes made in the cells are let out of the

cells to do their work outside. These are extracellular enzymes (‗extra' means ‗outside').
Examples:
1. Fungi and bacteria release extracellular enzymes in order to digest their food.
2. A mould growing on a piece of bread releases starch-digesting enzymes into the bread and absorbs sugars which the enzyme produces from the bread.
3. In the digestive system, extracellular enzymes are released into the stomach and intestines in order to digest the food.

Role of enzymes in the biological washing products
Biological washing powders contain enzymes (extracted from micro-organisms) such as proteases, lipases and amylases, often with high optimum temperatures, which help to break down the protein and fats stains, such as blood and egg, into smaller molecules. The small molecules are colourless and soluble in water, that can be washed away.

For example, the enzyme protease breaks down the colourful but insoluble protein molecules in the stains into simple amino acids. These are colourless and soluble simple molecules which can easily dissolve in water and be washed away. These powders are biodegradable and do not cause pollution.

Role of enzymes in seed germination
Before the germination of a seed, it is dry and contains non-active enzymes and stored food which is in the form of complex molecules and are not used by the seed. When seed is watered, it begins to germinate and absorb water. When sufficient water is absorbed, hydrolysis enzymes (or hydrolases) present in the seeds are activated. These enzymes break down (by hydrolysis) the food stored in the seed and convert it to small and soluble molecules which are transported to the growing parts of the plants and used in the growth of the seedling.

Role of enzymes in food industry
Food manufacturers often use enzymes for example, when juice is squeezed out of apples to make a drink, an enzyme called pectinase is usually added. Pectinase is an enzyme that breaks down the substance that holds the cell wall of the apple cell together. This makes it easier to squeeze most of the substances that make apple juice cloudy, turning it to a cler liquid.

Another enzyme that is often used is lactase. This is an enzyme that breaks down the sugar found in milk called lactose into another sugar called glucose. If lactase is added to the milk it breaks down all the lactose and it is safer for the people to drink who do not have lactase in their digestive system

56. ORGANISMS (COMMON TO ALL LIVING THINGS)

Organisms have the potential to carry out the life processes of nutrition, movement, growth, reproduction, respiration, sensitivity and excretion

The following characteristics are those that most biologists accept as being common to all living things. It is true that they may not always be displayed but even the most inactive of organisms has the potential to carry out all these functions. It is equally true that there are times in the life cycles of some organisms where all these functions appear to be suspended as is the case with seed producing organisms (Lotus seeds have been grown after being stored for 160 years).
Movement

Living things move in a directed and controlled way, moving of their own accord. Non-living things only move if they are pushed or pulled by something else. The majority of animals usually move their whole bodies often supported by specialized organs such as fins, wings and legs. These are called locomotors organs moving the animal from place to place.

Plant movement is not locomotors and does not generally involve moving the whole body. Leaves turning towards the light or shoots growing upwards whatever the orientation of the rest of the plant are examples of how plants move. These movements are generally very slow and not always obvious.

Growth

Living things grow. Most animals grow until they reach maturity and then remain at a constant size while plants usually continue to increase in size throughout their life span. It is important to recognize that growth is a permanent increase in measurable features such as volume, mass and length. Cells increase in number by dividing in a process called mitosis (making genetically exact copies). As the soft tissues increase, so there will be associated increase in size of skeletal support tissue such as bone, shell and wood.

When maturity is reached in animal's cell division continues only at a level to maintain consistent size and to repair loss through damage. Putting on weight as a result of over-eating is not considered to be biological growth in this context

Reproduction

Living things are able to reproduce themselves. If organisms fail to do this, populations will diminish and disappear as their members die from old age, disease, accidents, predation, etc. It is a fundamental law of biology that other living things can only produce living things; every living organism owes its existence to the reproductive activities of other organisms.

This is contrary to the misconceived ideas of spontaneous generation, which some people held in the past. The notion that cockroaches were formed out of crumbs on the bakery floor, that mould was formed out of decaying bread and that rotting sacks of grain turned into mice are examples of how spontaneous generation was thought to operate. Today, these ideas are discredited but they still often provide the stimulus for works of dramatic fiction.

Respiration

Living things respire. Respiration is a complex sequence of chemical reactions, which result in the release of energy from food. There are two types of respiratory process.

Aerobic respiration

Carried out by the vast majority of organisms, this involves oxygen. The by-products of the reaction are water and carbon dioxide both of which are eliminated as waste products. Oxygen is obtained from the air or water using organs designed to optimize gaseous exchange. These include the stomata in plants (small, size regulated pores), spiracles in arthropods, gills in fish and lungs in mammals. The uptake of oxygen and simultaneous elimination of carbon dioxide and water is commonly referred to as breathing. It is important to distinguish between breathing and respiration. It is tempting; particularly with younger children to use the well used term breathing as an all-embracing description of the respiratory process. However, this is not correct and could lead to the reinforcement of misconceptions.

Anaerobic respiration

When oxygen levels are at a low level, it is possible for some simpler organisms and parts of more complex ones to release energy from food without oxygen. This is a far less efficient process but a necessary alternative in some cases. The by-products of anaerobic respiration are different to aerobic. In humans, oxygen starved muscle cells will respire anaerobically under stress such as heavy physical activity. The by-product of this is lactic acid and it is this that causes the puffed out feeling. Yeast cells respire anaerobically in sugar solution producing alcohol as the by-product

Sensitivity

Living things are sensitive to their environment. This means that they detect and respond to events in the world around them. Simple uni-cellular organisms such as Amoeba have limited sensitivity, while higher organisms such as mammals are more sensitive and can react to very small changes in light, sound, touch, taste, smell, temperature, etc.

In higher animals specific organs are developed for the purpose of detecting stimuli. The organization of light sensitive cells into eyes of varying complexity from one species to another is an example.

Plants do not have sensory organs as such but there are clearly certain regions of their bodies such as the shoot tip that are sensitive to light, gravity, water and various chemicals.

Excretion

Living things excrete. Excretion is the removal from the body of waste products which result from normal life processes. Waste products such as carbon dioxide must be removed. If they are allowed to accumulate they cause poisoning which slows down vital chemical reactions. When excretory organs such as kidneys are damaged, the organism quickly displays symptoms of poisoning and death is rapid unless treated.

Excretion should not be confused with egestion, which is the removal from the body of substances with no food value that have passed unused through the digestive systems.

Feeding

Living things feed. Food is the material from which organisms through respiration obtain the energy required to sustain life and carry out all the other defining functions of living things. Food also provides the raw materials for growth and repair. The study of food and feeding is called nutrition.

There are two types of nutrition:

Autotrophic organisms make their own food by a process called photosynthesis. Green plants, for example, manufacture sugar and starch from carbon dioxide and water using the energy of sunlight to drive the necessary chemical reactions

Heterotrophic nutrition

Heterotrophic organisms obtain their food from the bodies of other organisms. This

is done in various ways.

Herbivores such as cattle, tortoises and sparrows eat plants.
Carnivores such as lions, crocodiles, sharks and kestrels eat the flesh of other animals.
Omnivores such as humans can eat both plants and animals.
Saprophytes such as many types of fungi and bacteria, obtain their food in liquid form from the remains of dead organisms. This feeding manifests itself as the process called decay.
Parasites such as tapeworms and mosquitoes live on or in another living organism (called the host) from which they obtain food.

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Keep remember me in your prayers….!

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