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SCHOOL OF SCIENCE AND HUMANITIES

DEPARTMENT OF PHYSICS

UNIT – I - Wave Mechanics – SPH1218

I. Introduction

Compton's formula established that an electromagnetic wave can behave like a particle of light when interacting with matter. In 1924, Louis de Broglie proposed a new speculative hypothesis that electrons and other particles of matter can behave like waves. Today, this idea is known as de Broglie's hypothesis of matter waves. In 1926, De Broglie's hypothesis, together with Bohr's early quantum theory, led to the development of a new theory of wave quantum mechanics to describe the physics of atoms and subatomic particles. Quantum mechanics has paved the way for new engineering inventions and technologies, such as the laser and magnetic resonance imaging (MRI). These new technologies drive discoveries in other sciences such as biology and chemistry.

According to de Broglie's hypothesis, massless photons as well as massive particles must satisfy one common set of relations that connect the energy *E* with the frequency *f*, and the linear momentum *p* with the wavelength

$$E = hf$$

$$\lambda = \frac{h}{p}.$$

Here, E and p are, respectively, the relativistic energy and the momentum of a particle. De Broglie's relations are usually expressed in terms of the wave vector $\vec{k}$, $k = 2\pi / \lambda$, and the wave frequency $\omega = 2\pi f$, as we usually do for waves:

$$E = \hbar\omega$$

$$\vec{p} = \hbar\,\vec{k}\ .$$

Wave theory tells us that a wave carries its energy with the group velocity. For matter waves, this group velocity is the velocity u of the particle. Identifying the energy E and momentum p of a particle with its relativistic energy mc^2 and its relativistic momentum mu, respectively, it follows from de Broglie relations that matter waves satisfy the following relation:

$$\lambda f = \frac{\omega}{k} = \frac{E/\hbar}{p/\hbar} = \frac{E}{p} = \frac{mc^2}{mu} = \frac{c^2}{u} = \frac{c}{\beta}$$

Using the concept of the electron matter wave, de Broglie provided a rationale for the quantization of the electron's angular momentum in the hydrogen atom, which was postulated in Bohr's quantum theory. The physical explanation for the first Bohr quantization condition comes naturally when we assume that an electron in a hydrogen atom behaves not like a particle but like a wave. To see it clearly, imagine a stretched guitar string that is clamped at both ends and vibrates in one of its normal modes. If the length of the string is l ((Figure)), the wavelengths of these vibrations cannot be arbitrary but must be such that an integer k number of half-wavelengths $\lambda/2$ fit exactly on the distance l between the ends. This is the condition $l = k\lambda/2$ for a standing wave on a string. Now suppose that instead of having the string clamped at the walls, we bend its length into a circle and fasten its ends to each other. This produces a circular string that vibrates in normal modes, satisfying the same standing-wave condition, but the number of half-wavelengths must now be an even number k, $k = 2n$, and the length l is now connected to the radius r_n of the circle. This means that the radii are not arbitrary but must satisfy the following standing-wave condition:

$$2\pi r_n = 2n\frac{\lambda}{2}.$$

Phase velocity

The **phase velocity** of a wave is the rate at which the phase of the wave propagates in space. This is the velocity at which the phase of any one frequency component of the wave travels. For such a component, any given phase of the wave (for example, the crest) will appear to travel at the phase velocity. The phase velocity is given in terms of the wavelength λ (lambda) and time period T.

Group velocity

The **group velocity** of a wave is the velocity with which the overall envelope shape of the wave's amplitudes.

The **phase velocity** is the ratio of the angular frequency to the wave number. The **group velocity** is the derivative of the angular frequency with respect to the wave number.

Davisson and Germer Experiment

The experimental setup for the Davisson and Germer experiment is enclosed within a vacuum chamber. Thus the deflection and scattering of electrons by the medium are prevented. The main parts of the experimental setup are as follows:

- Electron gun: An electron gun is a Tungsten filament that emits electrons via thermionic emission i.e. it emits electrons when heated to a particular temperature.
- Electrostatic particle accelerator: Two opposite charged plates (positive and negative plate) are used to accelerate the electrons at a known potential.
- Collimator: The accelerator is enclosed within a cylinder that has a narrow passage for the electrons along its axis. Its function is to render a narrow and straight (collimated) beam of electrons ready for acceleration.
- Target: The target is a Nickel crystal. The electron beam is fired normally on the Nickel crystal. The crystal is placed such that it can be rotated about a fixed axis.
- Detector: A detector is used to capture the scattered electrons from the Ni crystal. The detector can be moved in a semicircular arc as shown in the diagram above.

Browse more Topics under Dual Nature of Radiation and Matter

- Electron Emission
- Experimental Study of Photoelectric Effect
- Wave Nature of Matter
- Einstein's Photoelectric Equation: Energy Quantum of

Radiation The Thought Behind the Experimental Setup

The basic thought behind the Davisson and Germer experiment was that the waves reflected from two different atomic layers of a Ni crystal will have a fixed phase difference. After reflection, these waves will interfere either constructively or destructively. Hence producing a diffraction pattern. In the Davisson and Germer experiment waves were used in place of electrons. These electrons formed a diffraction pattern. The dual nature of matter was thus verified. We can relate the de Broglie equation and the Bragg's law as shown below:

From the de Broglie equation, we have:

$$\lambda = h/p$$

$$= h/\sqrt{2mE}$$

$$= h/\sqrt{2meV} \qquad \ldots (1)$$

where, m is the mass of an electron, e is the charge on an electron and h is the Plank's constant. Therefore for a given V, an electron will have a wavelength given by equation (1).

The following equation gives Bragg's Law:

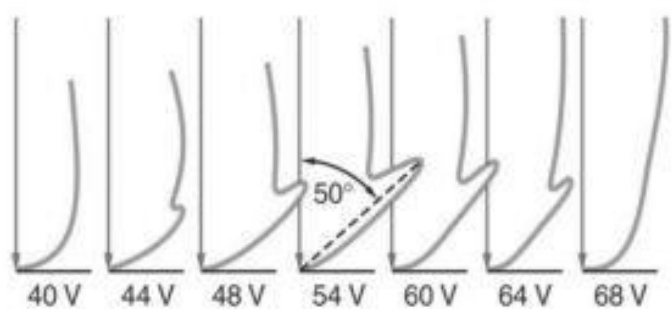
$$n\lambda = 2d \sin(90^{0} - \theta/2) \quad \ldots(2)$$

Since the value of d was already known from the X-ray diffraction experiments. Hence for various values of θ, we can find the wavelength of the waves producing a diffraction pattern from equation (2).

Observations of the Davisson and Germer Experiment

The detector used here can only detect the presence of an electron in the form of a particle. As a result, the detector receives the electrons in the form of an electronic current. The intensity (strength) of this electronic current received by the detector and the scattering angle is studied. We call this current as the electron intensity.

The intensity of the scattered electrons is not continuous. It shows a maximum and a minimum value corresponding to the maxima and the minima of a diffraction pattern produced by X-rays. It is studied from various angles of scattering and potential difference. For a particular voltage (54V, say) the maximum scattering happens at a fixed angle only as shown below:



Plots between I – *the intensity of scattering (X-axis) and the angle of scattering θ for given values of Potential difference.*

Results of the Davisson and Germer Experiment

From the Davisson and Germer experiment, we get a value for the scattering angle θ and a corresponding value of the potential difference V at which the scattering of electrons is maximum. Thus these two values from the data collected by Davisson and Germer, when used in equation (1) and (2) give the same values for λ. Therefore, this establishes the de Broglie's wave-particle duality and verifies his equation as shown below:

From (1), we have:

$$\lambda = h/\sqrt{2meV}$$

For V = 54 V, we have

$$\lambda = 12.27/\sqrt{54} = 0.167 \text{ nm} \quad \ldots. (3)$$

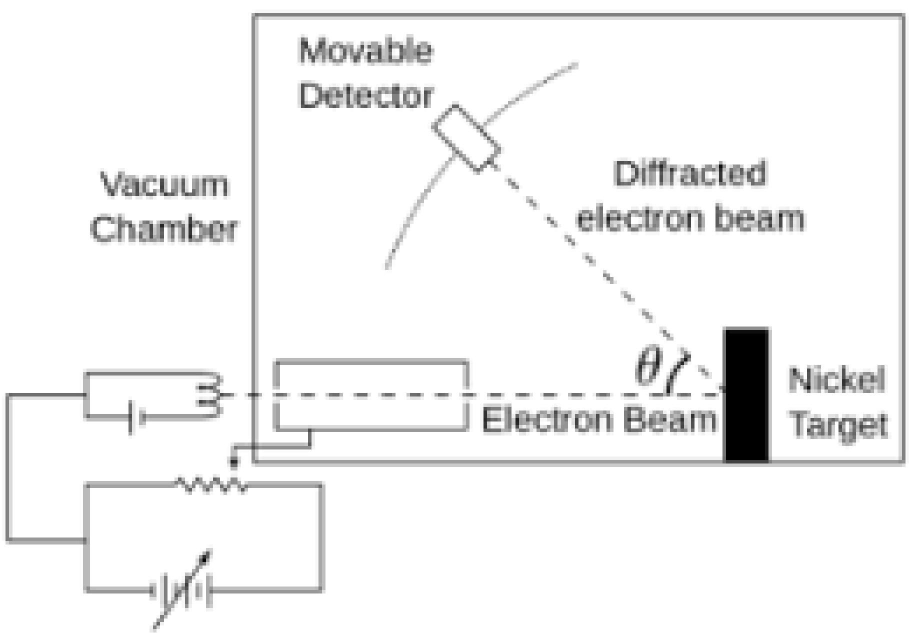
Now the value of 'd' from X-ray scattering is 0.092 nm. Therefore for V = 54 V, the angle of scattering is 50^{0}, using this in equation (2), we have:

$n\lambda = 2\ (0.092\ \text{nm})\sin(90^{0}-50^{0}/2)$

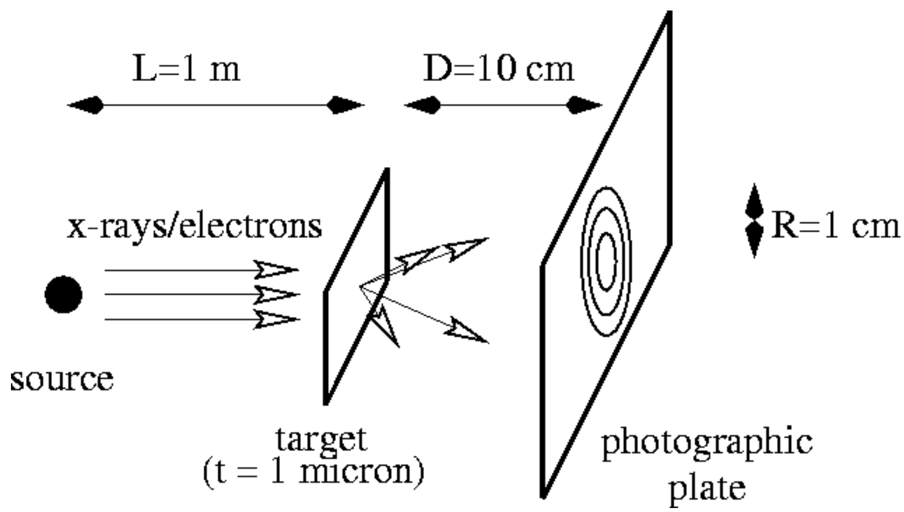
For n = 1, we have:

$\lambda = 0.165$ nm ….. (4)

Therefore the experimental results are in a close agreement with the theoretical values got from the de Broglie equation. The equations (3) and (4) verify the de Broglie equation.



G P Thomson Experiment



Electrons from an electron source were accelerated towards a positive electrode into which a small hole was drilled. The resulting narrow beam of electrons was directed towards a thin, rolled foil of gold. After passing through the hole in the gold foil, the electron beam was received on a photographic plate placed perpendicular to the direction of the beam. The diffraction pattern was in the form of continuous, alternate black and white rings as diffraction was due to the crystalline grains which were randomly oriented at all possible angles in the gold foil.

Electrons were scattered at different angles from the atoms of crystallites and produced interference pattern with maxima corresponding to those angles satisfying the Bragg condition. In terms of the probabilistic interpretation of matter waves, the probability of finding an electron scattered at an angle is exactly equal to computed intensity pattern of interfering waves associated with electron beam.

The diffraction pattern due to poly crystalline material was similar to the powder diffraction pattern of X-rays having wavelength equal to the de Broglie wavelength of electrons. The wavelength of electrons was varied by changing the incident energy of the electrons, then diameters of the diffraction rings changed proportionately according to the Bragg’s equation.

Uncertainty principle, also called **Heisenberg uncertainty principle** or **indeterminacy principle**, statement, articulated (1927) by the German physicist Werner Heisenberg, that the position and the velocity of an object cannot both be measured exactly, at the same time, even in theory.

The uncertainty principle is alternatively expressed in terms of a particle's momentum and position. The momentum of a particle is equal to the product of its mass times its velocity. Thus, the product of the uncertainties in the momentum and the position of a particle equals $h/(4\pi)$ or more. The principle applies to other related (conjugate) pairs of observables, such as energy and time: the product of the uncertainty in an energy measurement and the uncertainty in the time interval during which the measurement is made also equals $h/(4\pi)$ or more. The same relation holds, for an unstable atom or nucleus, between the uncertainty in the quantity of energy radiated and the uncertainty in the lifetime of the unstable system as it makes a transition to a more stable state.

UNIT – II- Nuclear Physics – SPH1218

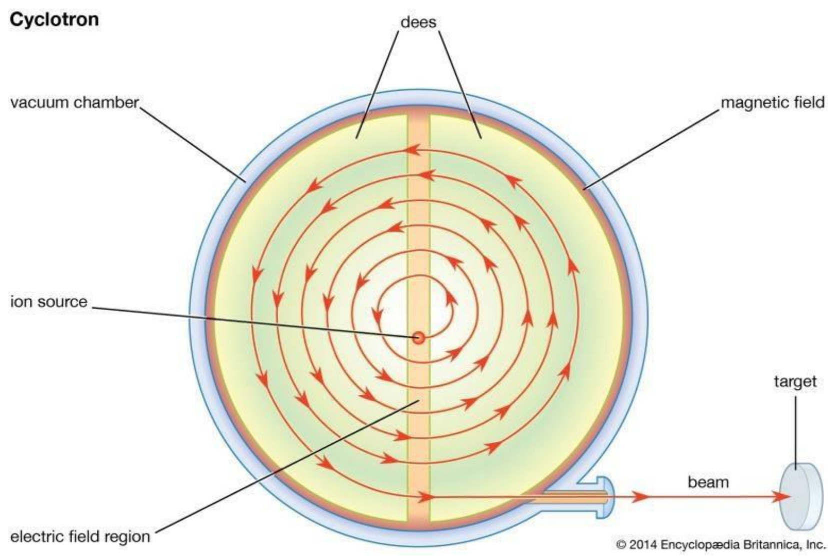
Particle accelerator

A particle accelerator is a machine that uses electromagnetic fields to propel charged particles to very high speeds and energies, and to contain them in well- defined beams. Smaller particle accelerators are used in a wide variety of applications, including particle therapy for oncological purposes, radioisotope production for medical diagnostics, ion implanters for manufacture of semiconductors, and accelerator mass spectrometers for measurements of rare isotopes such as radiocarbon. There are two basic classes of accelerators: electrostatic and electrodynamic (or electromagnetic) accelerators. Electrostatic accelerators use static electric fields to accelerate particles.

Cyclotron

A cyclotron is a type of particle accelerator invented by Ernest O. Lawrence in 1929–1930 at the University of California, Berkeley, and patented in 1932.A cyclotron accelerates charged particles outwards from the center along a spiral path. The particles are held to a spiral trajectory by a static magnetic field and accelerated by a rapidly varying (radio frequency) electric field. A cyclotron accelerates a charged particle beam using a high frequency alternating voltage which is applied between two hollow "D"-shaped sheet metal electrodes called "dees" inside a vacuum chamber. The dees are placed face to face with a narrow gap between them, creating a cylindrical space within them for the particles to move. The particles are injected into the center of this space. The dees are located between the poles of a large electromagnet which applies a static magnetic field *B* perpendicular to the electrode plane. The magnetic field causes the particles' path to bend in a circle due to the Lorentz force perpendicular to their direction of motion. If the particles' speeds were constant, they would travel in a circular path within the dees under the influence of the magnetic field. However a radio frequency (RF) alternating voltage of several thousand volts is applied between the dees. The voltage creates an oscillating electric field in the gap between the dees that accelerates the particles. The frequency is set so that the particles make one circuit during a single cycle of the voltage. To achieve this, the frequency must match the particle's cyclotron resonance frequency,

where B is the magnetic field strength, q is the electric charge of the particle and m is the relativistic mass of the charged particle. Each time after the particles pass to the other dee electrode the polarity of the RF voltage reverses. Therefore, each time the particles cross the gap from one dee electrode to the other, the electric field is in the correct direction to accelerate them. The particles' increasing speed due to these pushes causes them to move in a larger radius circle with each rotation, so the particles move in a spiral path outward from the center to the rim of the dees. When they reach the rim a small voltage on a metal plate deflects the beam so it exits the dees through a small gap between them, and hits a target located at the exit point at the rim of the chamber, or leaves the cyclotron through an evacuated beam tube to hit a remote target. Various materials may be used for the target, and the nuclear reactions due to the collisions will create secondary particles which may be guided outside of the cyclotron and into instruments for analysis.



Particle detector

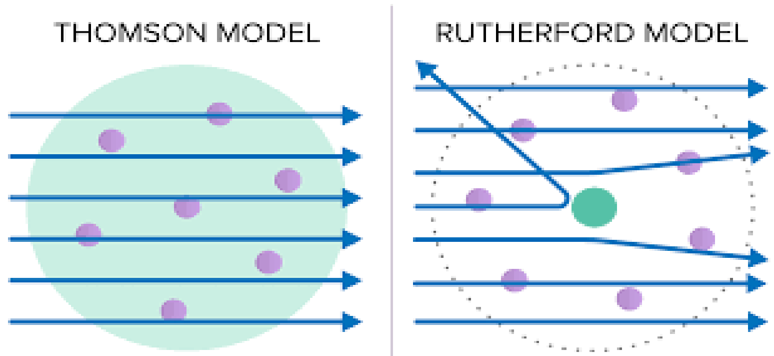
A particle detector, also known as a radiation detector, is a device used to detect, track, and/or identify ionizing particles, such as those produced by nuclear decay, cosmic radiation, or reactions in a particle accelerator. Detectors can measure the particle energy and other attributes such as momentum, spin, charge, particle type, in addition to merely registering the presence of the particle.

G M counter The Geiger–Miller tube or G–M tube is the sensing element of the Geiger counter instrument used for the detection of ionizing radiation

ARTIFICIAL TRANSMUTATION

Artificial transmutation is the conversion of one element into another by artificial means. It means, changing one element completely by bombarding it with some fundamental particles such as alpha particles, neutrons, deuterons etc. Artificial transmutation is formation of new element by artificial means.

The Gold Foil Experiment (Ernest Rutherford)



Rutherford began his graduate work by studying the effect of x-rays on various materials. Shortly after the discovery of radioactivity, he turned to the study of the - particles emitted by uranium metal and its compounds. Before he could study the effect of -particles on matter, Rutherford had to develop a way of counting individual -particles. He found that a screen coated with zinc sulfide emitted a flash of light each time it was hit by an –particle. They would then try to count the flashes of light given off by the ZnS screen. Rutherford found that a narrow beam of particles was broadened when it passed through a thin film of mica or metal. He therefore had Geiger measure the angle through which these -particles were scattered by a thin piece of metal foil. Because it is unusually ductile, gold can be made into a foil that is only 0.00004 cm thick. When this foil was bombarded with - particles, Geiger found that the scattering was small, on the order of one degree.

These results were consistent with Rutherford's expectations. He knew that the -particle had a considerable mass and moved quite rapidly. He therefore anticipated that virtually all of the - particles would be able to penetrate the metal foil, although they would be scattered slightly by collisions with the atoms through which they passed. In other words, Rutherford expected the - particles to pass through the metal foil the way a rifle bullet would penetrate a bag of sand.

Rutherford concluded that there was only one way to explain these results. He assumed that the positive charge and the mass of an atom are concentrated in a small fraction of the total volume and then derived mathematical equations for the scattering that would occur. These equations predicted that the number of -particles scattered through a given angle should be proportional to the thickness of the foil and the square of the charge on the nucleus, and inversely proportional to the velocity with which the -particles moved raised to the fourth power.

Q value for a reaction

The *Q* value for a reaction is the amount of energy absorbed or released during the nuclear reaction. The value relates to the enthalpy of a chemical reaction or the energy of radioactive decay products. It can be determined from the masses of reactants and products. *Q* values affect reaction rates. In general, the larger the positive *Q* value for the reaction, the faster the reaction proceeds, and the more likely the reaction is to "favor" the products.

$$Q = (M_r - M_p) \times 0.931 \text{ MeV}.$$

where the masses are in atomic mass units. Also both and are the sums of the reactant and product masses respectively.

Threshold Energy

The minimum energy that molecules need to have in order for a reaction to take place is called the threshold energy. Molecules that have a kinetic energy equal to or higher than the threshold energy will react.

Conservation of charge

Charge conservation is the principle that the total electric charge in an isolated system never changes. The net quantity of electric charge, the amount of positive charge minus the amount of negative charge in the universe, is always *conserved*. Charge conservation, considered as a physical conservation law, implies that the change in the amount of electric charge in any volume of space is exactly equal to the amount of charge flowing into the volume minus the amount of charge flowing out of the volume.

Law of Conservation of Nuclear Number

The law of Conservation of Nuclear Number states that the sum of protons and neutrons among species before and after a nuclear reaction will be the same.

Law of conservation of mass

The law of conservation of mass or principle of mass conservation states that for any system closed to all transfers of matter and energy, the mass of the system must remain constant over time, as the system's mass cannot change, so quantity can neither be added nor be removed. Therefore, the quantity of mass is conserved over time.

Law of Conservation of Energy

The law of conservation of energy states that energy can neither be created nor be destroyed. Although, it may be transformed from one form to another. If you take all forms of energy into account, the total energy of an isolated system always remains constant. All the forms of energy follow the law of conservation of energy. In brief, the law of conservation of energy states that in a closed system, i.e., a system that is isolated from its surroundings, the total energy of the system is conserved.

Conservation of parity

The principle that the parity of the total wave function describing a system of elementary particles is conserved. In fact it is not conserved in weak interactions.

Biological effects of radiation

Radiation is one of the best-investigated hazardous agents. Because of the vast accumulation of quantitative dose-response data, specialists are able to set environmental radiation levels so that applications of nuclear technologies may continue at a level of risk that is much less than with many other technologies.

The increased use of radioisotopes has led to increased concerns over the effects of these materials on biological systems (such as humans). All radioactive nuclides emit high- energy particles or electromagnetic waves. When this radiation encounters living cells, it can cause heating, break chemical bonds, or ionize molecules. The most serious biological damage results when these radioactive emissions fragment or ionize molecules. For example, alpha and beta particles emitted from nuclear decay reactions possess much higher energies than ordinary chemical bond energies. When these particles strike and penetrate matter, they produce ions and molecular fragments that are extremely reactive. The damage this does to biomolecules in living organisms can cause serious malfunctions in normal cell processes, taxing the organism's repair mechanisms and possibly causing illness or even death.

LONG-TERM SOMATIC EFFECTS Somatic cells are all cells of your body other than the reproductive cells. They can be damaged in a variety of ways, such as by chemical, biological and physical agents or by ionising radiation. The effects of the damage from ionising radiation can be short-term or long-term depending on the means and severity of the exposure. The most important long-term effect of radiation exposure is an increased chance of getting cancer.

RADIATION INDUCED CANCER IN HUMANS

There are many well-documented cases of radiation induced cancer in humans. The early scientists who worked with X- rays and radioactive substances did not realize the risk. Many died from skin and bone cancers and from leukemia.

Severe Mental Retardation

The developing human brain is very vulnerable to radiation damage between the 8th and 15th weeks of the pregnancy. The risk of severe mental retardation is high at 40%/Gy, with a threshold of a few hundred mGy. This is based on the Japanese data at high dose rates.

Reduced Intelligence I.Q. testing of Japanese born after the A-bomb explosions suggests a loss of 30 I.Q. points for those who were exposed to 1 Gy at 8 - 15 weeks after conception. The loss was smaller for weeks 16 - 25, and there was no evidence of any effect after the 26th week.

5. Childhood Cancer

There is disagreement between the Japanese data at high doses and large populations of children who received small doses of prenatal radiation for medical reasons. The estimated risk of childhood cancer and leukaemia range from 2% to 6% per sievert.

Control of radiation hazards

There are three factors that control the amount, or dose, of radiation received from a source. Radiation exposure can be managed by a combination of these factors:

1. Time: Reducing the time of an exposure reduces the effective dose proportionally.

An example of reducing radiation doses by reducing the time of exposures might be improving operator training to reduce the time they take to handle a radioactive source.

2. Distance: Increasing distance reduces dose due to the inverse square law. Distance can be as simple as handling a source with forceps rather than fingers.
3. Shielding: Sources of radiation can be shielded with solid or liquid material, which absorbs the energy of the radiation. The term 'biological shield' is used for absorbing material placed around a nuclear reactor, or other source of radiation, to reduce the radiation to a level safe for humans

a. Internal radiation hazard and its control—contamination control

An internal radiation hazard arises when radioactive material is taken into the body and its control is exercised by controlling the levels of contamination in the environment. The

possibility of contaminating the environment from the sources used in radioisotope instruments is extremely unlikely and this section is only included for completeness and for those who, for any reason, wish to manufacture their own sources.

Radiation Hazard

In general, there is no radiation hazard to shipboard personnel provided that the antenna is rotating. However, *harmful* effects, particularly to the eyes, can be experienced at very short distances from a stationary antenna if energy is being radiated.

Protection Control of Radiation Measures

In larger organizations, the control of radiation hazards is the responsibility of specialists known as *health physicists* or, sometimes, *health chemists*. Their main duty is to ensure that work is carried out without hazard to the health of the people involved.
The protection follows three stages: prevention, supervision, and after-control. Preventive measures include use of fume hoods, α-boxes, radiation shielding, tongs, etc., as discussed above. The supervision stage involves the use of radiation instruments to monitor the radiation level . Small TLD, film or pocket pen dosimeters are used for individual monitoring (§8.9). For

spills and contamination of hands, shoes, etc., special contamination instruments (counters) are used which are more sensitive than the monitoring dose instruments

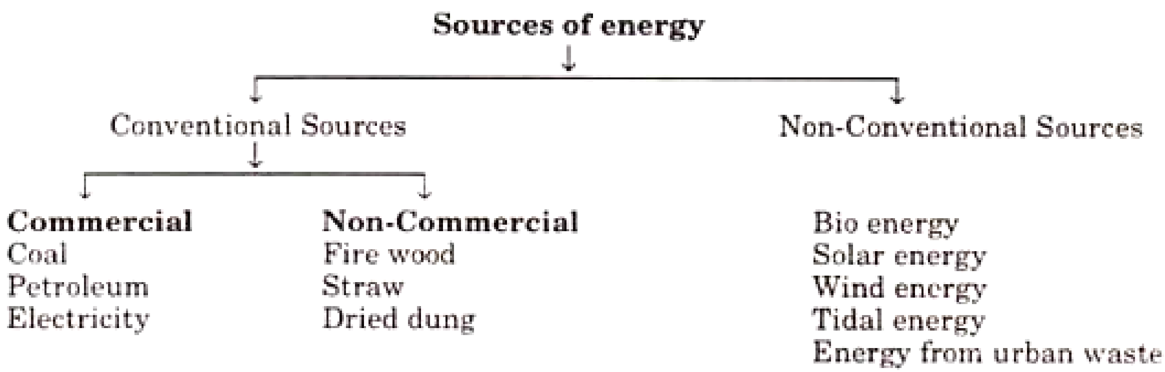
Fundamental to radiation protection is the avoidance or reduction of dose using the simple protective measures of time, distance and shielding. The duration of exposure should be limited to that necessary, the distance from the source of radiation should be maximised, and the source shielded wherever possible. To measure personal dose uptake in occupational or emergency exposure, for external radiation personal dosimeters are used, and for internal dose to due to ingestion of radioactive contamination, bioassay techniques are applied.

UNIT – III- Energy Physics – SPH1218

Conventional and Non-Conventional Sources of Energy

Two Main Sources of Energy

The sources of energy are of following types:



1. Conventional Sources of Energy:

These sources of energy are also called non-renewable sources. These sources of energy are in limited quantity except hydro-electric power.
These are coal, petroleum and electricity. These are called commercial energy because they have a price and consumer must pay the price to purchase them.

(a) Coal and Lignite:

Coal is the major source of energy. Coal deposits in India are 148790 million tonnes. Total lignite reserves found at Naively are 3300 million tonnes. In 1950-51, annual production of coal was 32 million tonnes. In 2005-06, annual production of coal was 343 million tonnes. Lignite production was 20.44 million tonnes in 2005-06. According to an estimate, coal reserves in India would last about 130 years. India is now the fourth largest coal producing country in the world. Coal deposits are mainly found in Orissa, Bihar, Bengal and Madhya Pradesh. It provides employment to 7 lakh workers.

(b) Oil and Natural Gas:

In these days oil is considered as the most important source of energy in India and the world. It is widely used in automobiles, trains, planes and ships etc. In India it is found in upper

Assam, Mumbai High and in Gujarat. The resources of oil are small in India.

In 1950-51, the total production of oil in India was 0.3 million tonnes. It increased to 32.4 million tonnes in 2000-01. Despite tremendous increase in oil production. India still imports 70% of has oil requirements from abroad. In 1951, there was only one oil refinery in Assam.

Natural gas has been the most important source of energy since last two decades. It can be produced in two ways:

(i) With petroleum products as associated gas.

(ii) Free gas obtained from gas fields in Assam, Gujarat and Andhra Pradesh.

It is used in fertilizer and petro-chemical plants and gas based thermal power plants.

Total production of natural gas was 31.96 billion cubic metre in 2003-04.

(c) Electricity:

Electricity is the common and popular source of energy. It is used in commercial and domestic purposes. It is used for lighting, cooking, air conditioning and working of electrical appliances like T.V., fridge and washing machine.

In 2000-01 agriculture sector consumed 26.8%, industrial sector 34.6% and 24% of electricity was used for domestic purposes and 7% was used for commercial purpose. Railways consumed 2.6% and miscellaneous consumption was 5.6%.

There are three main sources of power generation:

1. Thermal Power
2. Hydro-electric power
3. Nuclear Power

1. Thermal Power:

It is generated in India at various power stations with the help of coal and oil. It has been a major source of electric power. In 2004-05, its share in total installed capacity was 70 percent.

2. Hydroelectric Power:

It is produced by constructing dams over overflowing rivers. For example, Bhakra Nangal Project, Damodor Valley Project and Hirakund Project etc. In 1950-51, installed capacity of hydroelectricity was 587.4 MW and in 2004-05, it was 19600 MW.

3. Nuclear Power:

India has also developed nuclear power. Nuclear Power plants use uranium as fuel. This fuel is cheaper than coal. India has nuclear power plants at Tarapur, Kota (Rajasthan) Kalapakam (Chennai) Naroura (UP). Its supply accounts for only 3 percent of the total installed capacity.

2. Non-Conventional Sources of Energy:

Besides conventional sources of energy there are non-conventional sources of energy. These are also called renewable sources of energy. Examples are Bio energy, solar energy, wind energy and tidal energy. Govt. of India has established a separate department under the Ministry of Energy called as the Department of Non-conventional Energy Sources for effective exploitation of non-conventional energy.

The various sources are given below:

1. Solar Energy:

Energy produced through the sunlight is called solar energy. Under this programme, solar photovoltaic cells are exposed to sunlight and in the form of electricity is produced. Photovoltaic cells are those which convert sun light energy into electricity. In year 1999-2000, 975 villages were illuminated through solar energy. Under Solar Thermal Programme, solar energy is directly obtained. Sunlight is converted into thermal power. Solar energy is used for cooking, hot water and distillation of water etc.

2. Wind Energy:

This type of energy can be produced by harnessing wind power. It is used for operating water pumps for irrigation purposes. Approximately 2756 wind pumps were set up for this purpose. In seven states, wind power operated power houses were installed and their installed capacity was 1000 MW. India has second position in wind power energy generation.

3. Tidal Energy:

Energy produced by exploiting the tidal waves of the sea is called tidal energy. Due to the absence of cost-effective technology, this source has not yet been tapped.

4. Bio Energy:

This type of energy is obtained from organic matter.

It is of two kinds:

(i) Biogas:

Biogas is obtained from Gobar Gas Plant by putting cow dung into the plant. Besides producing gas this plant converts gobar into manure. It can be used for cooking, lighting and generation of electricity. 26.5 lakh biogas plants had been established by the year 2003-04. They produce more than 225 lakh tonnes of manure. About 1828 large community biogas plants have been established in the country.

(ii) Biomass:

It is also of a source of producing energy through plants and trees. The purpose of biomass programme is to encourage afforestation for energy. So that fuel for the generation of energy based on gas technique and fodder for the cattle could be obtained, 56 MW capacity for the generation of biomass energy has been installed.

5. Energy from Urban Waste:

Urban waste poses a big problem for its disposal. Now it can be used for generation of power. In Timarpur (Delhi) a power Ration of 3.75 capacity has been set up to generate energy from the garbage.

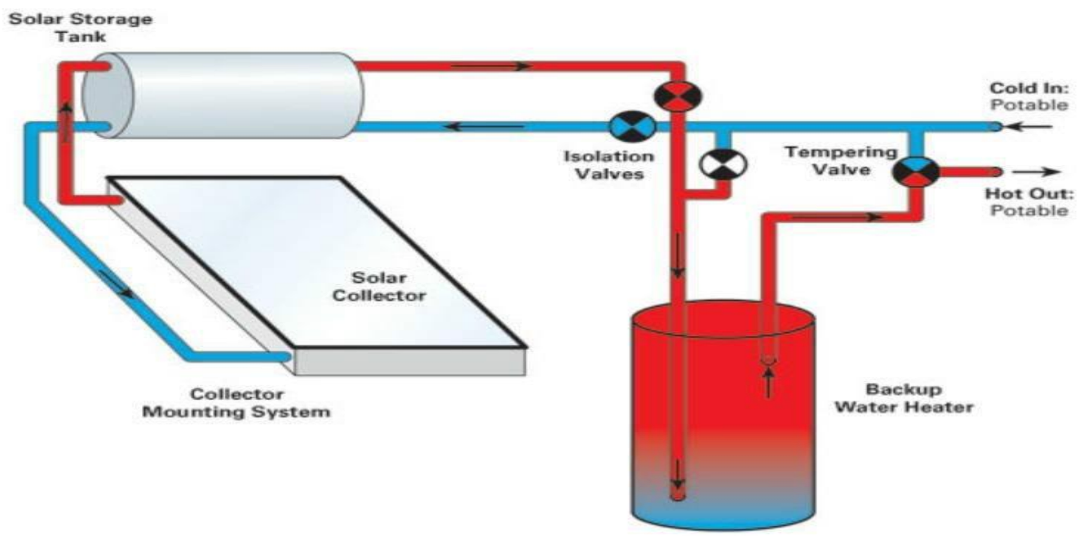
SOLAR RADIATION

Solar radiation is radiant energy emitted by the sun, particularly electromagnetic energy. About half of the radiation is in the visible short-wave part of the electromagnetic spectrum. The other half is mostly in the near-infrared part, with some in the ultraviolet part of the spectrum. Sun and earth geometry:

A Solar constant is defined, commonly taken as 1353 W/m2 , though there are some variations in the estimates. Further, distinct from the solar constant, solar radiation as received by a plane normal to sun's rays, if there is no attenuation in the atmosphere also is defined. This quantity varies from day to day due to ellipticity of the earth's orbit. SUN - EARTH SIZE, POSITION The geometry of the sun-earth relationships is shown schematically in Fig. 3.1. The distance between the sun and earth varies by 1.7 % due to the elliptic orbit of the of the earth around the sun. The mean earth distance is 1.495 x 1011 m. The sun subtends an angle of 32' with the earth. The diameter of the sun is 1.39 x 109 m and that of earth, 1.27 x 107 m. The distance between the earth and the sun determines the intensity of solar radiation that can be received by the earth, notwithstanding the atmospheric attenuation.

SOLAR WATER HEATER

The solar water heater absorbs light by means of a collector placed on the roof and converts it into heat. It passes this heat to a water tank by means of a circulating pump. This exchange is triggered by the thermal regulator, but only when the collector is hotter than the water in the tank. This prevents the circulating pumps using electricity needlessly. Conversely, it also prevents overheating.



Circulation Systems

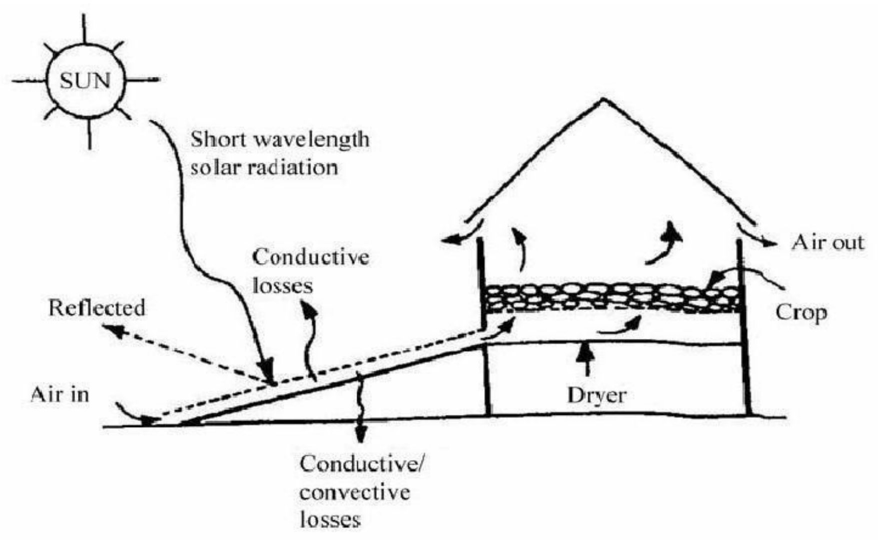
Direct systems circulate water through solar collectors where it is heated by the sun. The heated water is then stored in a tank, sent to a tankless water heater, or used directly.

Closed loop, or indirect, systems use a non-freezing liquid to transfer heat from the sun to water in a storage tank. The sun's thermal energy heats the fluid in the solar collectors. Then, this fluid passes through a heat exchanger in the storage tank, transferring the heat to the water. The non-freezing fluid then cycles back to the collectors.

Flat-plate collectors consist of copper tubes fitted to flat absorber plates. The most common configuration is a series of parallel tubes connected at each end by two pipes, the inlet and outlet manifolds. The flat plate assembly is contained within an insulated box and covered with tempered glass.

Evacuated tube collectors are the most efficient collectors available. A glass or metal tube containing the water or heat transfer fluid is surrounded by a larger glass tube. The space between them is a vacuum, so very little heat is lost from the fluid.

SOLAR DRYER



The solar dryer is a relatively simple concept. An important feature of solar drying devices is the size of the solar collectors. Depending on the quantity of goods to be dried, collectors must have the capacity to provide sufficient quantities of hot air to the drying chamber. Collectors which are too small in proportion to the amount of food to be dried will result in failed attempts and spoiled food.

The basic principles employed in a solar dryer are:

Converting light to heat: Any black on the inside of a solar dryer will improve the effectiveness of turning light into heat.

Trapping heat: Isolating the air inside the dryer from the air outside the dryer makes an important difference. Using a clear solid, like a plastic bag or a glass cover, will allow light to enter, but once the light is absorbed and converted to heat, a plastic bag or glass cover will trap the heat inside. This makes it possible to reach similar temperatures on cold and windy days as on hot days.

Moving the heat to the food. Both the natural convection dryer and the forced convection dryer use the convection of the heated air to move the heat to the food.

Types of Solar Dryer

A solar dryer can most conveniently be classified as direct or indirect.

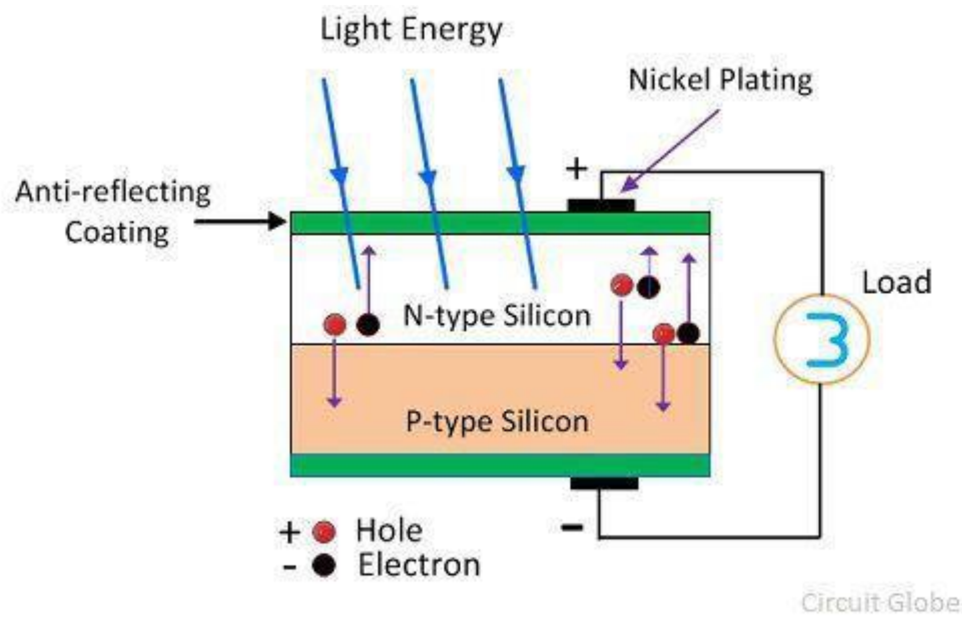
Direct Solar Dryer

The item to be dried is exposed directly to solar radiation through a transparent material that covers the structure. The heat generated from the solar energy is used to dry the crops or food items and also heats up the surroundings. The main disadvantage of using the direct mode is that the heat that will be absorbed by the item cannot be controlled. Available in many sizes, ranging from kilograms to metric tons, the simplicity of the product and its affordability are its USP. Indirect Solar Dryer

As the name suggests, this method does not expose the crop directly to the sunlight. The solar radiation is absorbed and converted into heat by another surface (like a black top) usually called the collector. Air that will be used for drying is passed over this surface and gets heated, which is then used to dry the food item inside the dryer. The main advantage of indirect mode of drying is that the temperatures can be controlled. The sizes can vary from kilograms to metric tons, but it is expensive and more complex to construct when compared to direct solar dryers.

SOLAR ENERGY- Conversion of light into electrical energy

Photovoltaic or Solar Cell



Definition: The Photovoltaic cell is the semiconductor device that converts the light into electrical energy. The voltage induces by the PV cell depends on the intensity of light incident on it. The name Photovoltaic is because of their voltage producing capability.

The semiconductor materials like arsenide, indium, cadmium, silicon, selenium and gallium are used for making the PV cells. Mostly silicon and selenium are used for making the cell. Consider the figure below shows the constructions of the silicon photovoltaic cell. The upper surface of the cell is made of the thin layer of the p-type material so that the light can easily enter into the material. The metal rings are placed around p-type and n-type material which acts as their positive and negative output terminals respectively.

Photovoltaic energy is the conversion of sunlight into electricity. A photovoltaic cell, commonly called a solar cell or PV, is the technology used to convert solar energy directly into electrical power. A photovoltaic cell is a nonmechanical device usually made from silicon alloys. Sunlight is composed of photons, or particles of solar energy. These photons contain various amounts of energy corresponding to the different wavelengths of the solar spectrum. When photons strike a photovoltaic cell, they may be reflected, pass right through, or be absorbed. Only the absorbed photons provide energy to generate electricity. When enough sunlight (energy) is absorbed by the material (a semiconductor), electrons are dislodged from the material's atoms. Special treatment of the material surface during manufacturing makes the front surface of the cell more receptive to free electrons, so the electrons naturally migrate to the surface. When the electrons leave their position, holes are formed. When many electrons, each carrying a negative charge, travel toward the front surface of the cell, the resulting imbalance of charge between the cell's front and back surfaces creates a voltage potential like the negative and positive terminals of a battery. When the two surfaces are connected through an external load, electricity flows. The photovoltaic cell is the basic building block of a photovoltaic.

Solar Power Advantages and Disadvantages

Advantages:

1. Solar power is pollution free and causes no greenhouse gases to be emitted after installation
2. Reduced dependence on foreign oil and fossil fuels
3. Renewable clean power that is available every day of the year, even cloudy days produce some power
4. Return on investment unlike paying for utility bills
5. Virtually no maintenance as solar panels last over 30 years
6. Creates jobs by employing solar panel manufacturers, solar installers, etc. and in turn helps the economy
7. Excess power can be sold back to the power company if grid intertied
8. Ability to live grid free if all power generated provides enough for the home / building
9. Can be installed virtually anywhere, in a field to on a building
10. Use batteries to store extra power for use at night
11. Solar can be used to heat water, power homes and building, even power cars
12. Safer than traditional electric current

13. Efficiency is always improving so the same size solar that is available today will become more efficient tomorrow

Disadvantages

1. High initial costs for material and installation and long ROI
2. Needs lots of space as efficiency is not 100% yet
3. No solar power at night so there is a need for a large battery bank
4. Some people think they are ugly
5. Devices that run on DC power directly are more expensive
6. Depending on geographical location the size of the solar panels vary for the same power generation
7. Cloudy days do not produce much energy
8. Solar panels are not being massed produced due to lack of material and technology to lower the cost enough to be more affordable
9. Solar powered cars do not have the same speeds and power as typical gas-powered cars
10. Lower production in the winter month

UNIT – IV- Crystallography and Fiber Optics – SPH1218

Crystal Systems: They are 7 Crystal Systems. The structures of all crystals can be classified according to the symmetry of the unit cells. There are in total 7 groups, collectively called Crystal Systems: Tricinic, Monoclinic, Orthorhombic, Tetragonal, Trigonal, Hexagonal, and Cubic

In crystallography, crystal structure is a description of the ordered arrangement of atoms, ions or molecules in a crystalline material. Ordered structures occur from the intrinsic nature of the constituent particles to form symmetric patterns that repeat along the principal directions of three-dimensional space in matter.

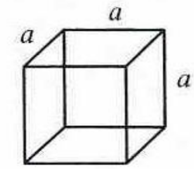
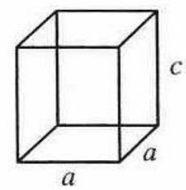
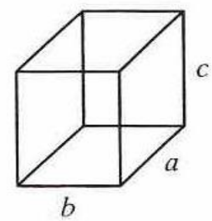
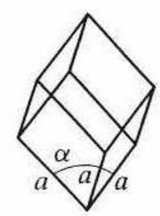
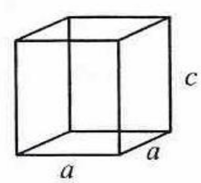
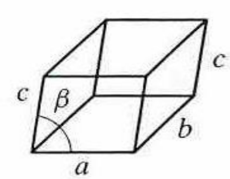
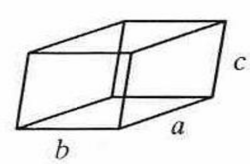
The smallest group of particles in the material that constitutes this repeating pattern is the unit cell of the structure. The unit cell completely reflects the symmetry and structure of the entire crystal, which is built up by repetitive translation of the unit cell along its principal axes.

Different type of crystals

There are seven crystal lattice systems.

1. Cubic or Isometric: These are not always cube-shaped. You'll also find octahedrons (eight faces) and dodecahedrons (10 faces).
2. Tetragonal: Similar to cubic crystals, but longer along one axis than the other, these crystals forming double pyramids and prisms.
3. Orthorhombic: Like tetragonal crystals except not square in cross-section (when viewing the crystal on end), these crystals form rhombic prisms or dipyramids (two pyramids stuck together).
4. Hexagonal: When you look at the crystal on end, the cross-section is a six-sided prism or hexagon.
5. Trigonal: These crystals possess a single 3-fold axis of rotation instead of the 6- fold axis of the hexagonal division.
6. Triclinic: These crystals are not usually symmetrical from one side to the other, which can lead to some fairly strange shapes.
7. Monoclinic: Like skewed tetragonal crystals, these crystals often form prisms and double pyramids.

This is a very simplified view of crystal structures. In addition, the lattices can be primitive (only one lattice point per unit cell) or non-primitive (more than one lattice point per unit cell). Combining the 7 crystal systems with the 2 lattice types yields the 14 Bravais Lattices (named after Auguste Bravais)

System	Axial lengths and angles[a]	Unit cell geometry
Cubic	$a = b = c, \alpha = \beta = \gamma = 90°$	
Tetragonal	$a = b \neq c, \alpha = \beta = \gamma = 90°$	
Orthorhombic	$a \neq b \neq c, \alpha = \beta = \gamma = 90°$	
Rhombohedral	$a = b = c, \alpha = \beta = \gamma \neq 90°$	
Hexagonal	$a = b \neq c, \alpha = \beta = 90°, \gamma = 120°$	
Monoclinic	$a \neq b \neq c, \alpha = \gamma = 90° \neq \beta$	
Triclinic	$a \neq b \neq c, \alpha \neq \beta \neq \gamma \neq 90°$	

Type of Crystals by Properties

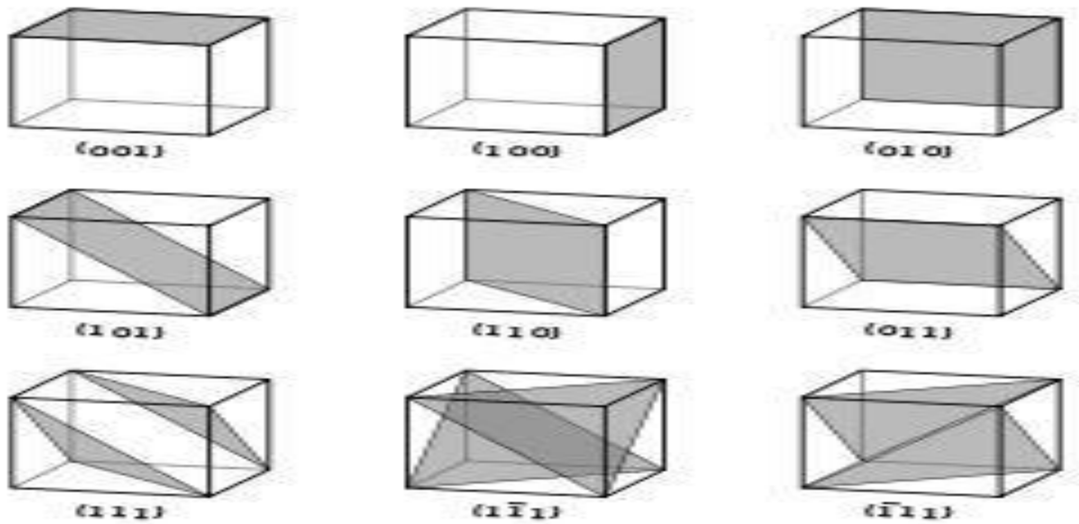
There are four main categories of crystals, as grouped by their chemical and physical properties.

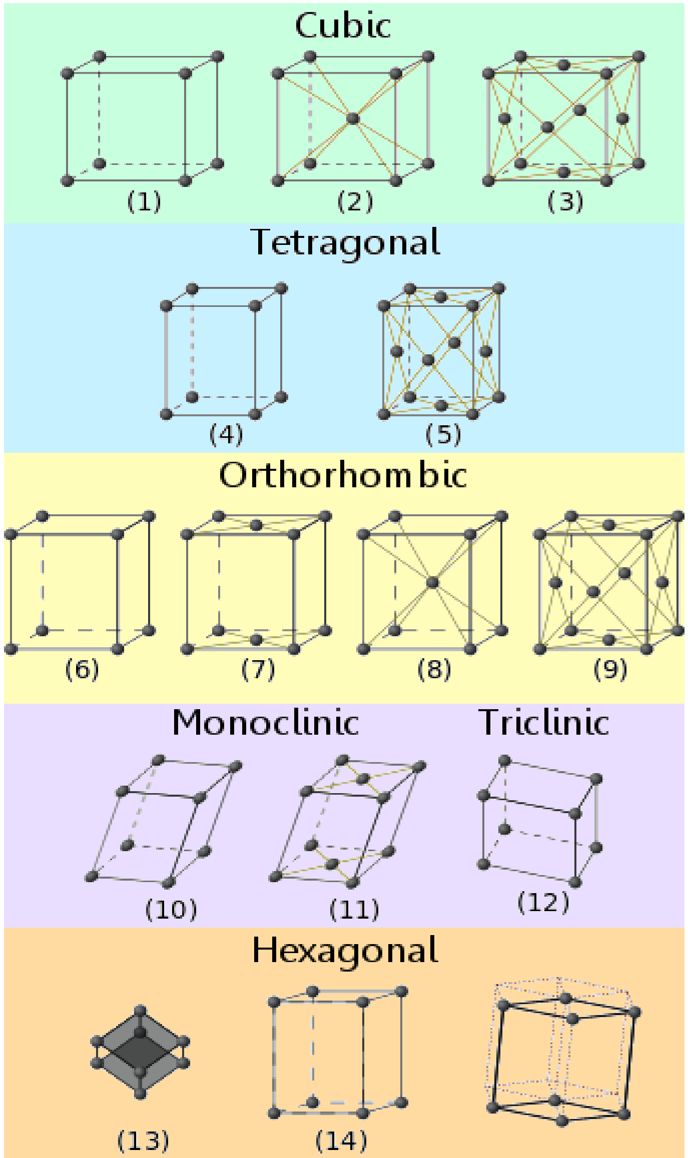
1. Covalent Crystals: A covalent crystal has true covalent bonds between all of the atoms in the crystal. You can think of a covalent crystal as one big molecule. Many covalent crystals have extremely high melting points. Examples of covalent crystals include diamond and zinc sulfide crystals.
2. Metallic Crystals: Individual metal atoms of metallic crystals sit on lattice sites. This leaves the outer electrons of these atoms free to float around the lattice. Metallic crystals tend to be very dense and have high melting points.
3. Ionic Crystals: The atoms of ionic crystals are held together by electrostatic forces (ionic bonds). Ionic crystals are hard and have relatively high melting points. Table salt (NaCl) is an example of this type of crystal.
4. Molecular Crystals: These crystals contain recognizable molecules within their structures. A molecular crystal is held together by non-covalent interactions, like van der Waals forces or hydrogen bonding. Molecular crystals tend to be soft with relatively low melting points. Rock candy, the crystalline form of table sugar or sucrose, is an example of a molecular crystal.

Miller indices

Miller indices are used to specify directions and planes. • These directions and planes could be in lattices or in crystals. • The number of indices will match with the dimension of the lattice or the crystal. • E.g. in 1D there will be 1 index and 2D there will be two indices etc.

Vectors and planes in a crystal lattice are described by the three-value Miller index notation. This syntax uses the indices ℓ, m, and n as directional parameters. By definition, the syntax (ℓmn) denotes a plane that intercepts the three points a_1/ℓ, a_2/m, and a_3/n, or some multiple thereof. That is, the Miller indices are proportional to the inverses of the intercepts of the plane with the unit cell (in the basis of the lattice vectors). If one or more of the indices is zero, it means that the planes do not intersect that axis (i.e., the intercept is "at infinity"). A plane containing a coordinate axis is translated so that it no longer contains that axis before its Miller indices are determined. The Miller indices for a plane are integers with no common factors. Negative indices are indicated with horizontal bars, as in (123). In an orthogonal coordinate system for a cubic cell, the Miller indices of a plane are the Cartesian components of a vector normal to the plane.





The simplest and most symmetric, the cubic or isometric system, has the symmetry of a cube, that is, it exhibits four threefold rotational axes oriented at 109.5° (the tetrahedral angle) with respect to each other. These threefold axes lie along the body diagonals of the cube. The other six lattice systems, are hexagonal, tetragonal, rhombohedral (often confused with the trigonal crystal system), orthorhombic, monoclinic and triclinic.

Bravais lattices

Bravais lattices, also referred to as space lattices, describe the geometric arrangement of the lattice points, and therefore the translational symmetry of the crystal. The three dimensions of space afford 14 distinct Bravais lattices describing the translational symmetry. All crystalline materials recognized today, not including quasicrystals, fit in one of these arrangements. The fourteen three-dimensional lattices, classified by lattice system, are shown above.

The crystal structure consists of the same group of atoms, the *basis*, positioned around each and every lattice point. This group of atoms therefore repeats indefinitely in three dimensions according to the arrangement of one of the Bravais lattices. The characteristic rotation and mirror symmetries of the unit cell is described by its crystallographic point group.

Crystal systems

A crystal system is a set of point groups in which the point groups themselves and their corresponding space groups are assigned to a lattice system. Of the 32 point groups that exist in three dimensions, most are assigned to only one lattice system, in which case the crystal system and lattice system both have the same name. However, five point groups are assigned to two lattice systems, rhombohedral and hexagonal, because both lattice systems exhibit threefold rotational symmetry. These point groups are assigned to the trigonal crystal system.

Atomic packing factor

In crystallography, atomic packing factor (APF), packing efficiency or packing fraction is the fraction of volume in a crystal structure that is occupied by constituent particles. It is a dimensionless quantity and always less than unity. In atomic systems, by convention, the APF is determined by assuming that atoms are rigid spheres. The radius of the spheres is taken to be the maximum value such that the atoms do not overlap. For one- component crystals (those that contain only one type of particle), the packing fraction is represented mathematically by

APF =Volume of atom in one unit cell/ volume of the unit cell

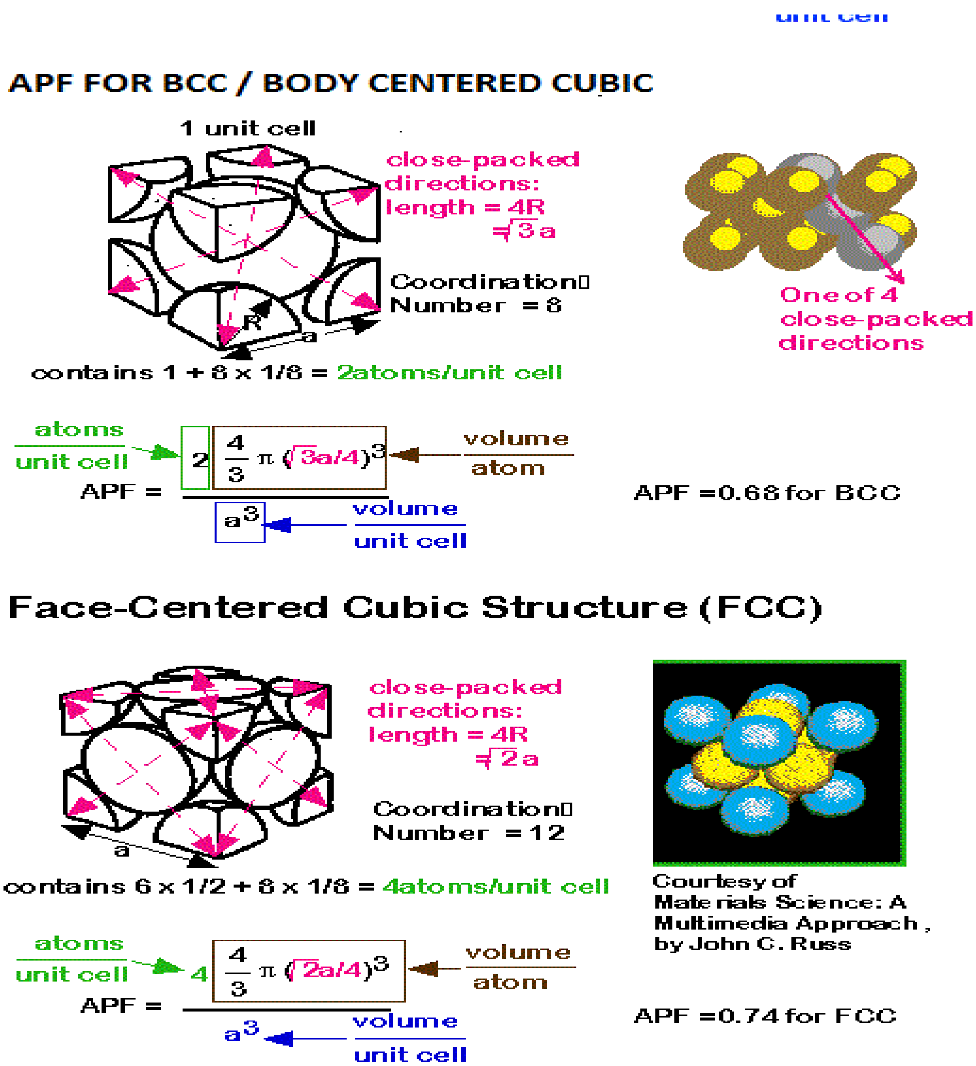
Simple cubic

For a simple cubic packing, the number of atoms per unit cell is one. The side of the unit cell is of length 2*r*, where *r* is the radius of the atom.

Calculation part APF= 52.36 percentage

Body-cantered cubic

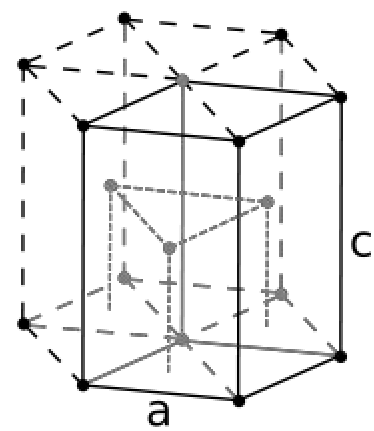
The primitive unit cell for the body-centered cubic crystal structure contains several fractions taken from nine atoms (if the particles in the crystal are atoms): one on each corner of the cube and one atom in the center. Because the volume of each of the eight corner atoms is



shared between eight adjacent cells, each BCC cell contains the equivalent volume of two atoms (one central and one on the corner).

Each corner atom touches the center atom. A line that is drawn from one corner of the cube through the center and to the other corner passes through 4*r*, where *r* is the radius of an atom. By geometry, the length of the diagonal is $a\sqrt{3}$. Therefore, the length of each side of the BCC structure can be related to the radius of the atom by Side a = 4r/root 3. APF= 68.1 percentage.

Hexagonal close-packed

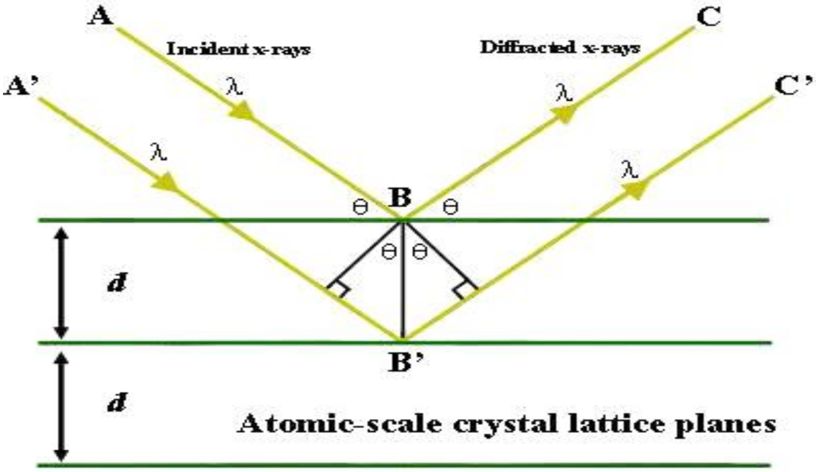


For the hexagonal close-packed structure the derivation is similar. Here the unit cell (equivalent to 3 primitive unit cells) is a hexagonal prism containing six atoms (if the particles in the crystal are atoms). Indeed, three are the atoms in the middle layer (inside the prism); in addition, for the top and bottom layers (on the bases of the prism)

BRAGG'S LAW

When a crystal is bombarded with X-rays of a fixed wavelength (similar to spacing of the atomic-scale crystal lattice planes) and at certain incident angles, intense reflected X-rays are produced when the wavelengths of the scattered X-rays interfere constructively. In order for the waves to interfere constructively, the differences in the travel path must be equal to integer multiples of the wavelength. When this constructive interference occurs, a diffracted beam of X-rays will leave the crystal at an angle equal to that of the incident beam.

To illustrate this feature, consider a crystal with crystal lattice planar distances d (right). Where the travel path length difference between the ray paths ABC and A'B'C' is an integer multiple of the wavelength, constructive interference will occur for a combination of that specific wavelength, crystal lattice planar spacing and angle of incidence (Θ). Each rational plane of atoms in a crystal will undergo refraction at a single, unique angle (for X-rays of a fixed wavelength).



The general relationship between the wavelength of the incident X-rays, angle of incidence and spacing between the crystal lattice planes of atoms is known as Bragg's Law, expressed as:

$$n \lambda = 2d \sin\Theta$$

where n (an integer) is the "order" of reflection, λ is the wavelength of the incident X-rays, d is the interplanar spacing of the crystal and Θ is the angle of incidence.

Types of bonding in crystals

Atoms bond during chemical reactions to result in crystal formation. When the electrons of atoms combine with surrounding atoms, a chemical bond is consummated, and crystals are formed.

Ionic Bonds

When ionic crystals are formed, electrons jump their orbits to bond with the corresponding supporting atom. The resultant combination of negatively or positively charged electrostatic forces stabilizes ions.

Covalent Bonds

A covalent bond, as the name suggests, is a crystal structure in which the electrons do not leave their orbits. Electrons, instead, are shared between two atoms. A shared electron in this way binds every two adjacent atoms. The bound atoms further share another electron from the atoms next to them and so on. Covalent bonding between the atoms of a substance results in the formation of a geometrical crystal.

Van der Waals Bonds

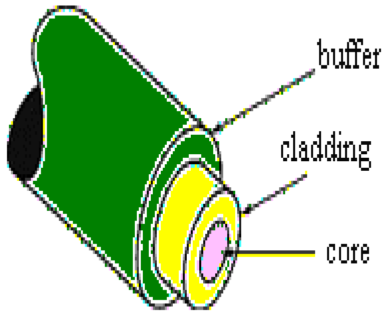
A Van der Waals bond is a weak interaction between the atoms of a substance, resulting in soft-consistency crystals. The outer orbit of the atoms is completely filled with shared electrons, but their charge keeps transferring.

Hydrogen Bonds

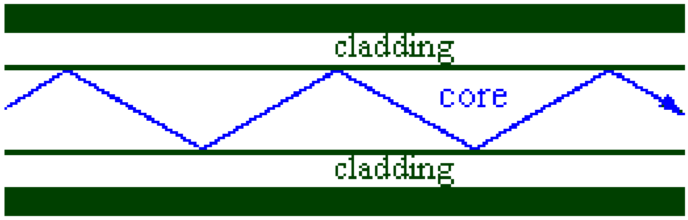
A hydrogen bond is formed when an atom of hydrogen is attracted toward the respective electrons of the corresponding atoms. This interferes with the crystal formation. A hydrogen atom, after being bound to another atom, is pulled toward the negative charge of a neighboring molecule.

OPTICAL FIBER

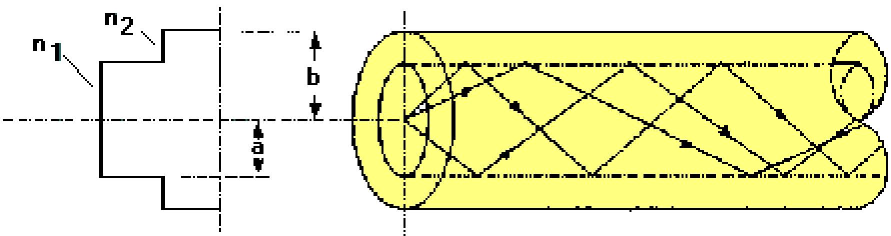
Light can be guided by planar or rectangular wave guides, or by optical fibers. An optical fiber consists of three concentric elements, the core, the cladding and the outer coating, often called the buffer. The core is usually made of glass or plastic. The core is the light-carrying portion of the fiber. The cladding surrounds the core. The cladding is made of a material with a slightly lower index of refraction than the core. This difference in the indices causes total internal reflection to occur at the core-cladding boundary along the length of the fiber. Light is transmitted down the fiber and does not escape through the sides of the fiber.



- Fiber Optic Core:
  - the inner light-carrying member with a high index of refraction.
- Cladding:
  - the middle layer, which serves to confine the light to the core. It has a lower index of refraction.
- Buffer:
  - the outer layer, which serves as a "shock absorber" to protect the core and cladding from damage. The coating usually comprises one or more coats of a plastic material to protect the fiber from the physical environment. Sometimes metallic sheaths are added to the coating for further physical protection.

o 

- Light injected into the fiber optic core and striking the core-to-cladding interface at an angle greater than the critical angle is reflected back into the core. Since the angles of incidence and reflection are equal, the light ray continues to zigzag down the length of the fiber. The light is trapped within the core. Light striking the interface at less than the critical angle passes into the cladding and is lost.



MULTIMODE STEP-INDEX FIBER

- Fibers for which the refractive index of the core is a constant and the index changes abruptly at the core-cladding interface are called step-index fibers. Step-index fibers are available with core diameters of 100 mm to 1000 mm. They are well suited to applications requiring high-power densities, such as delivering laser power for medical and industrial applications.

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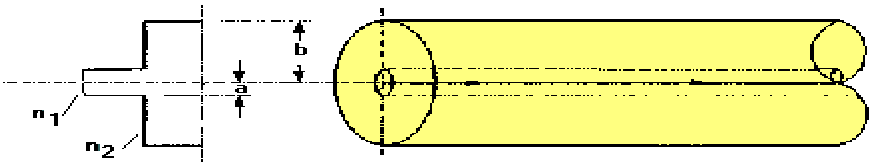


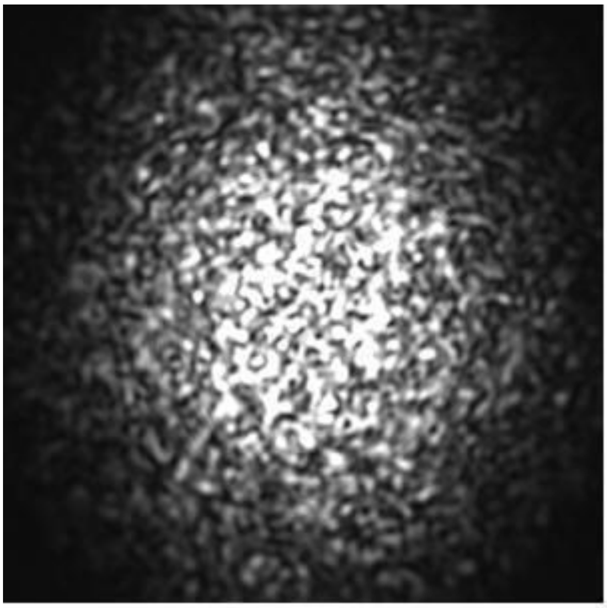
SINGLE MODE STEP-INDEX FIBER

Multimode step-index fibers trap light with many different entrance angles, each mode in a step-index multimode fiber is associated with a different entrance angle. Each mode therefore travels along a different path through the fiber. Different propagating modes have different velocities. As an optical pulse travels down a multimode fiber, the pulse begins to spread. Pulses that enter well separated from each other will eventually overlap each other. This limits the distance over which the fiber can transport data. Multimode step-index fibers are not well suited for data transport and communications.

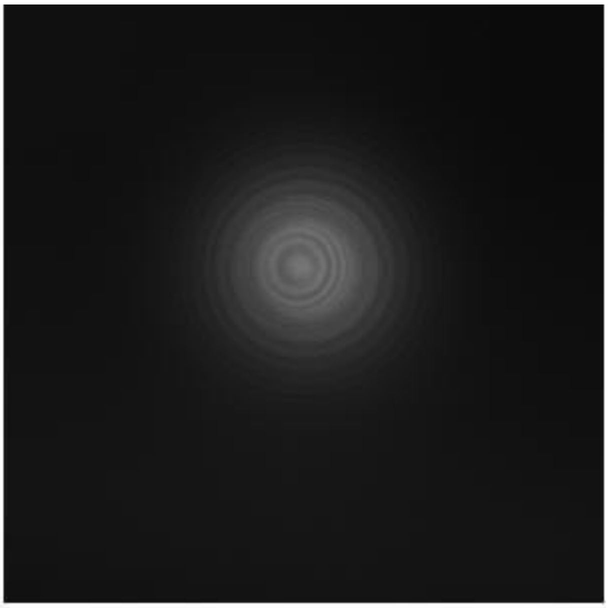
In a multimode graded-index fiber the core has an index of refraction that decreases as the radial distance from the center of the core increases. As a result, the light travels faster near the edge of the core than near the center. Different modes therefore travel in curved paths with nearly equal travel times. This greatly reduces the spreading of optical pulses.

A single mode fiber only allows light to propagate down its center and there are no longer different velocities for different modes. A single mode fiber is much thinner than a multimode fiber and can no longer be analyzed using geometrical optics. Typical core diameters are between 5 mm and 10 mm.

When laser light is coupled into a fiber, the distribution of the light emerging from the other end reveals if the fiber is a multimode or single mode fiber.



Light emerging from a multi-mode fiber



Light emerging from a single-mode fiber

Optical fibers are used widely in the medical field for diagnoses and treatment. Optical fibers can be bundled into flexible strands, which can be inserted into blood vessels, lungs and other parts of the body. An Endoscope is a medical tool carrying two bundles of optic fibers inside one long tube. One bundle directs light at the tissue being tested, while the other bundle carries light reflected from the tissue, producing a detailed image. Endoscopes can be designed to look at regions of the human body, such as the knees, or other joints in the body

Need of fiber optic communication

Fiber optic communication system has emerged as most important communication system. Compared to traditional system because of following requirements :

1. In long haul transmission system, there is need of low loss transmission medium
2. There is need of compact and least weight transmitters and receivers.
3. There is need of increase span of transmission.
4. There is need of increased bit rate-distance product.

A fiber optic communication system fulfills these requirements, hence most widely accepted.

Optical Fiber Communication System

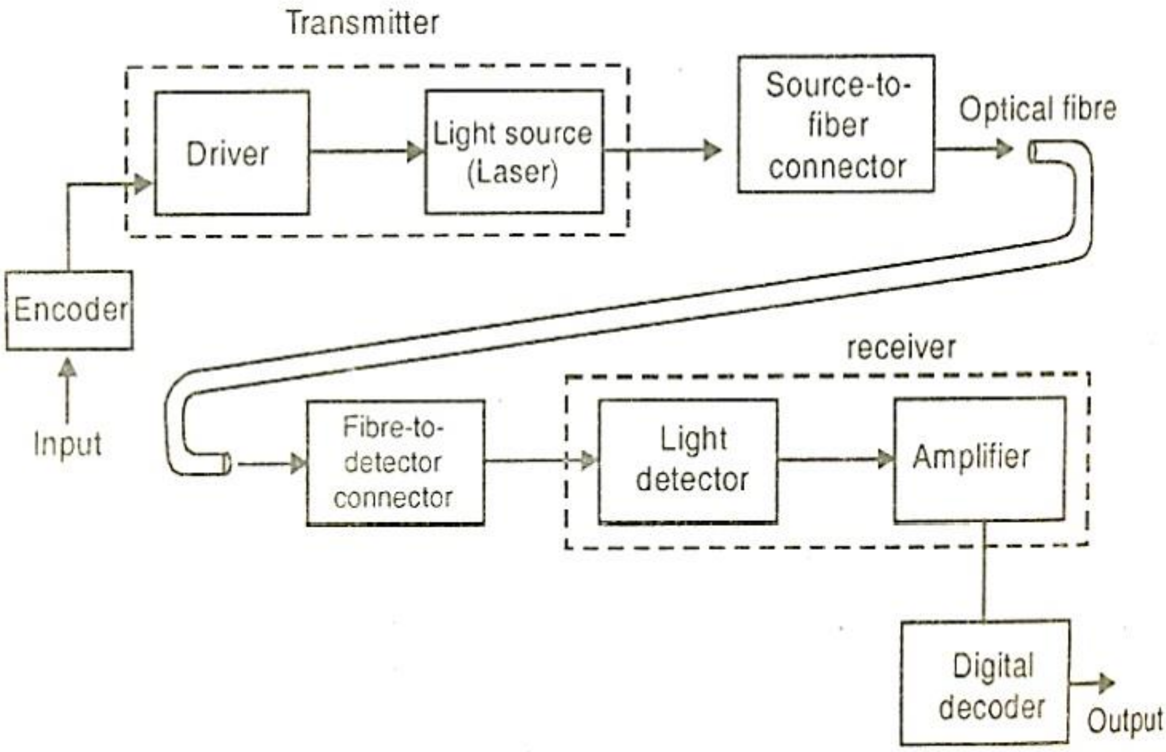
Basic block diagram of optical fiber communication system consists of following important blocks.

The major components of an optical fiber communication system are

i. The optical transmitter
ii. The optical fiber
iii. The optical receiver

PRINCIPLE:

Basically, a fiber optic system converts an electrical signal to an infrared light signal. This signal is transmitted through an optical fiber. At the end of the optical fiber, it is reconverted into an electric signal



Working:

1. Encoder encodes the information in the binary sequence zeros and ones.

a. Encoder is an electric circuit where in the information is encoded into binary sequences of zeros and one. In the light wave transmitter each ‘one’ corresponds to an electrical pulse and ‘zero’ corresponds to an absence of a pulse. These electrical pulses are used to turn a light source on and off very rapidly. The driver converts the incoming electrical signal into a form that will operate with the light source.

2. These electrical pulses are used to turn a light source on and off rapidly.

3. The optical fiber acts as a wave guide and transmits the optical pulses towards the receiver, by the principle of total internal reflection.

4. The light detector receives the optical pulses and converts them into electrical pulses. These signals are amplified by the amplifier.

5. The amplified signals are decoded by the decoder.

Advantages of Optical Fiber Communications

1. Wide bandwidth

The light wave occupies the frequency range between 2 x 1012 Hz to 3.7 x 1012 Hz. Thus the information carrying capability of fiber optic cables is much higher.

2. Low losses

Fiber optic cables offers bery less signal attenuation over long distances. Typically it is less than 1 dB/km. This enables longer distance between repeaters.

3. Immune to cross talk

Fiber optic cables has very high immunity to electricaland magnetic field. Since fiber optic cables are non-conductors of electricity hence they do not produce magnetic field. Thus fiber optic cables are immune to cross talk between cables cause dby magnetic induction.

4. Interference immune

Fiber optic cable immune to conductive and radiative interferences caused by electrical noise sources such as lighting, electric motors, fluorescent lights.

5. Light weight

As fiber cables are made of silica glass or plastic which is much lighter than copper or aluminium cables. Light weight fiber cables are cheaper to transport.

6. Small size

The diameter of fiber is much smaller compared to other cables, therefore fiber calbe is small in size, requires less storage space.

7. More strength

Fiber cables are stronger and rugged hence can support more weight.

8. Security

Fiber cables are more secure than other cables. It is almost impossible to tap into a fiber cable as they do n ot radiate signals.

No ground loops exist between optical fibers hence they are more secure.

9. Long distance transmission

Because of less attenuation transmission at a longer distance is possible.

10. Environment immune

Fiber cables are more immune to environmental extremes. They can operate over a large temperature variations. Also they are not affected by corrosive liquids and gases.

11. Sage and easy installation

Fiber cables are safer and easier to install and maintain. They are non- conductors hence there is no shock hazards as no current or voltage is associated with them. Their small size and light weight feature makes installation easier.

12. Less cost

Cost of fiber optic system is less compared to any other system.

Disadvantages of Optical Fiber Communications

1. High initial cost

The initial cost of installation or setting up cost is very high compared to all other system.

2. Maintenance and repairing cost

The maintenance and repairing of fiber optic systems is not only difficult but expensive also.

3. Jointing and test procedures

Since optical fibers are of very small size. The fiber joining process is very costly and requires skilled manpower.

4. Tensile stress

Optical fibers are more susceptible to buckling, bending and tensile stress than copper cables. This leades to restricted practice to use optical fiber technology to premises and floor backbones with a few interfaces to the copper cables.

5. Short links

Even though optical fiber cables are inexpensive, it is still not cost effective to replace every small conventional connector (e.g. between computers and peripherals), as the price of optoelectronic transducers are very high.

6. Fiber losses

The amount of optical fiber available to the photodetector at the end of fiber length depends on various fiber losses such as scattering, dispersion, attenuation and reflection.

Applications of Optical Fiber Communications

Applications of optical fiber communications include telecommunications, data communications, video control and protection switching, sensors and power applications.

1. Telephone networks

Optical waveguide has low attenuation, high transmission bandwidth compared to copper lines, therefore numbers of long haul co-axial trunks links between telephone exchanges are being replaced by optical fiber links.

2. Urban broadband service networks

Optical waveguide provides much larger bandwidth than co-axial cable, also the number of repeaters required is reduced considerably.

Modern suburban communications involves videotext, videoconferencing videotelephony, switched broadband communication network. All these can be supplied over a single fiber optic link. Fiber optic cables is the solution to many of today’s high speed, high bandwidth data communication problems and will continue to play a large role in future telecom and data-com networks.

UNIT – V- Electronics – SPH1218

A Zener diode is a silicon semiconductor device that permits current to flow in either a forward or reverse direction. The diode consists of a special, heavily doped p-n junction, designed to conduct in the reverse direction when a certain specified voltage is reached. The Zener diode has a well-defined reverse-breakdown voltage, at which it starts conducting current, and continues operating continuously in the reverse-bias mode without getting damaged. Additionally, the voltage drop across the diode remains constant over a wide range of voltages, a feature that makes Zener diodes suitable for use in voltage regulation.

Working Principle of Zener Diode

When a PN junction diode is reverse biased, the depletion layer becomes wider. If this reverse biased voltage across the diode is increased continually, the depletion layer becomes more and more wider. At the same time, there will be a constant reverse saturation current due to minority carriers.

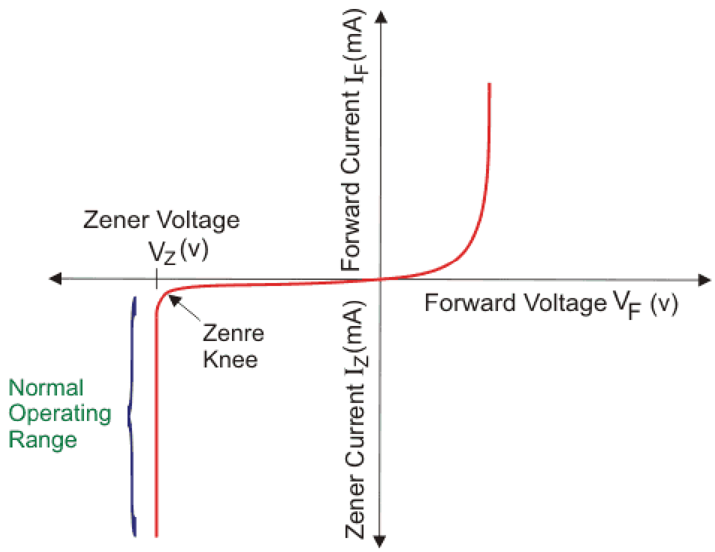
Zener diode operation

The Zener diode operates just like the normal diode when in the forward-bias mode, and has a turn-on voltage of between 0.3 and 0.7 V. However, when connected in the reverse mode, which is usual in most of its applications, a small leakage current may flow. As the reverse voltage increases to the predetermined breakdown voltage (Vz), a current starts flowing through the diode. The current increases to a maximum, which is determined by the series resistor, after which it stabilizes and remains constant over a wide range of applied voltage.

Zener Diode Circuit

Zener Diode is nothing but a single diode connected in a reverse bias, we have already stated that. A diode connected in reverse bias position in a circuit is shown below, Characteristics of a Zener Diode

Now, discussing about the diode circuits we should look through the graphical representation of the operation of the **zener diode**. Normally, it is called the V-I characteristics of a Zener diode.



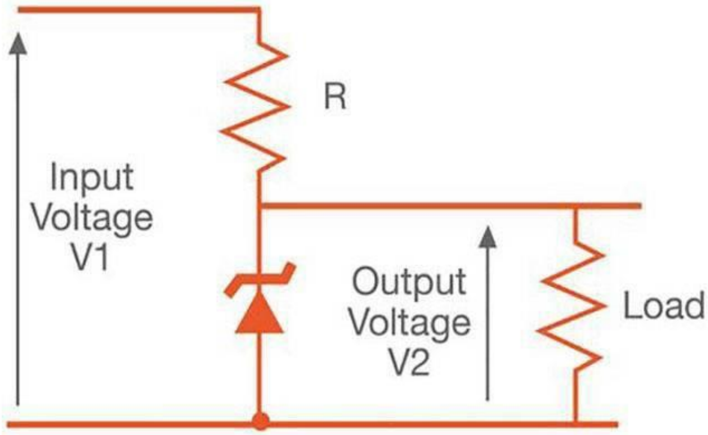
The above diagram shows the V-I characteristics of a zener diode. When the diode is connected in forward bias, this diode acts as a normal diode but when the reverse bias voltage is greater than zener voltage, a sharp breakdown takes place. In the V-I characteristics above V_Z is the zener voltage. It is also the knee voltage because at this point the current increases very rapidly.

Zener diode applications

Zener diodes are used for voltage regulation, as reference elements, surge suppressors, and in switching applications and clipper circuits.

Voltage regulator

The load voltage equals breakdown voltage VZ of the diode. The series resistor limits the current through the diode and drops the excess voltage when the diode is conducting.



Zener Diodes can be used to produce a stabilised voltage output with low ripple under varying load current conditions. By passing a small current through the diode from a voltage source, via a suitable current limiting resistor (RS), the zener diode will conduct sufficient current to maintain a voltage drop of Vout.

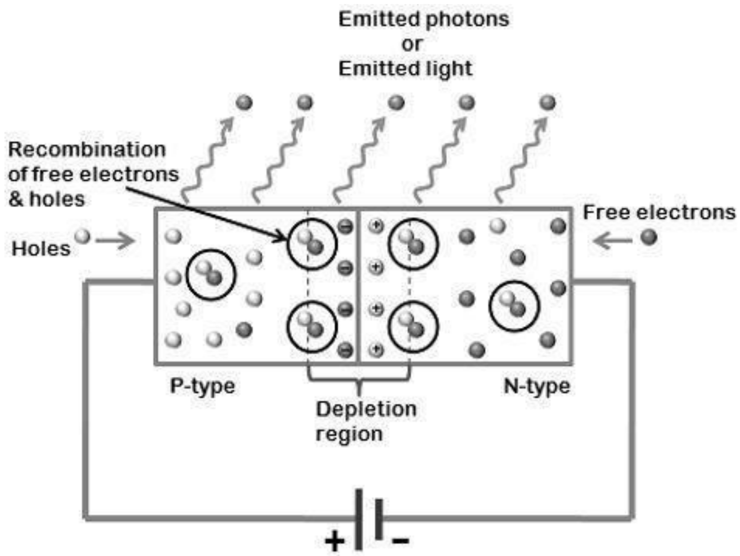
We remember from the previous tutorials that the DC output voltage from the half or full- wave rectifiers contains ripple superimposed onto the DC voltage and that as the load value changes so to does the average output voltage. By connecting a simple zener stabiliser circuit as shown below across the output of the rectifier, a more stable output voltage can be produced.

Light Emitting Diode- LED

An LED is just like a normal p n junction diode, but with light-emitting properties. Its construction and working can be explained as follows.

Working of LED

Like an ordinary diode, the LED diode works when it is forward biased. In this case, the n-type semiconductor is heavily doped than the p-type forming the p-n junction. When it is forward biased, the potential barrier gets reduced and the electrons and holes combine at the depletion layer (or active layer), light or photons are emitted or radiated in all directions. A typical figure blow showing light emission due electron-hole pair combining on forward biasing.



Definition: The LED is a PN-junction diode which emits light when an electric current passes through it in the forward direction. In the LED, the recombination of charge carrier takes place. The electron from the N-side and the hole from the P-side are combined and gives the energy in the form of heat and light. The LED is made of semiconductor material which is colourless, and the light is radiated through the junction of the diode.

Working of LED

The working of the LED depends on the quantum theory. The quantum theory states that when the energy of electrons decreases from the higher level to lower level, it emits energy in the form of photons. The energy of the photons is equal to the gap between the higher and lower level. The LED is connected in the forward biased, which allows the current to flows in the forward direction. The flow of current is because of the movement of electrons in the opposite direction. The recombination shows that the electrons move from the conduction band to valence band and they emits electromagnetic energy in the form of photons. The energy of photons is equal to the gap between the valence and the conduction band.

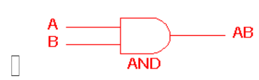
DIGITAL GATE CIRCUITS

Boolean functions may be practically implemented by using electronic gates. The following points are important to understand. Electronic gates require a power supply.

Gate **INPUTS** are driven by voltages having two nominal values, e.g. 0V and 5V.representing logic 0and logic 1 respectively. The **OUTPUT** of a gate provides two nominal values of voltage only, e.g. 0V and 5V representing logic 0 and logic 1 respectively. In general, there is only one output to a logic gate except in some special cases.

There is always a time delay between an input being applied and the output responding. Digital systems are said to be constructed by using logic gates. These gates are the AND, OR, NOT, NAND, NOR, EXOR and EXNOR gates. The basic operations are described below with the aid of truth tables.

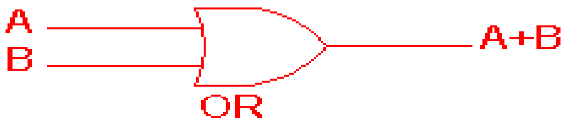
AND gate



A	B	A.B
0	0	0
0	1	0
1	0	0
1	1	1

The AND gate is an electronic circuit that gives a **high** output (1) only if **all** its inputs are high. A dot (.) is used to show the AND operation i.e. A.B. Bear in mind that this dot is sometimes omitted i.e. AB

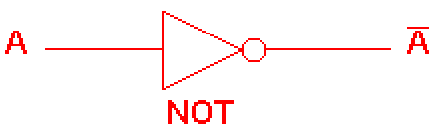
OR gate



2 Input OR gate		
A	B	A+B
0	0	0
0	1	1
1	0	1
1	1	1

The OR gate is an electronic circuit that gives a high output (1) if **one or more** of its inputs are high. A plus (+) is used to show the OR operation.

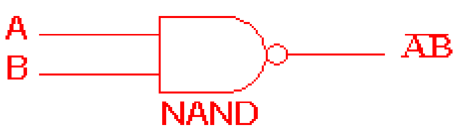
NOT gate



NOT gate	
A	$\bar{A}$
0	1
1	0

The NOT gate is an electronic circuit that produces an inverted version of the input at output. It is also known as an *inverter*. If the input variable is A, the inverted output is known as NOT A. This is also shown as A', or A with a bar over the top, as shown at the outputs. The diagrams below show two ways that the NAND logic gate can be configured toproduce a NOT gate. It can also be done using NOR logic gates in the same way.

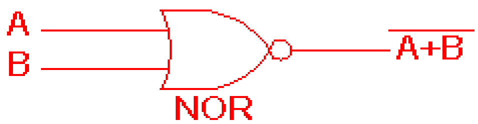
NAND gate



2 Input NAND gate		
A	B	$\overline{A.B}$
0	0	1
0	1	1
1	0	1
1	1	0

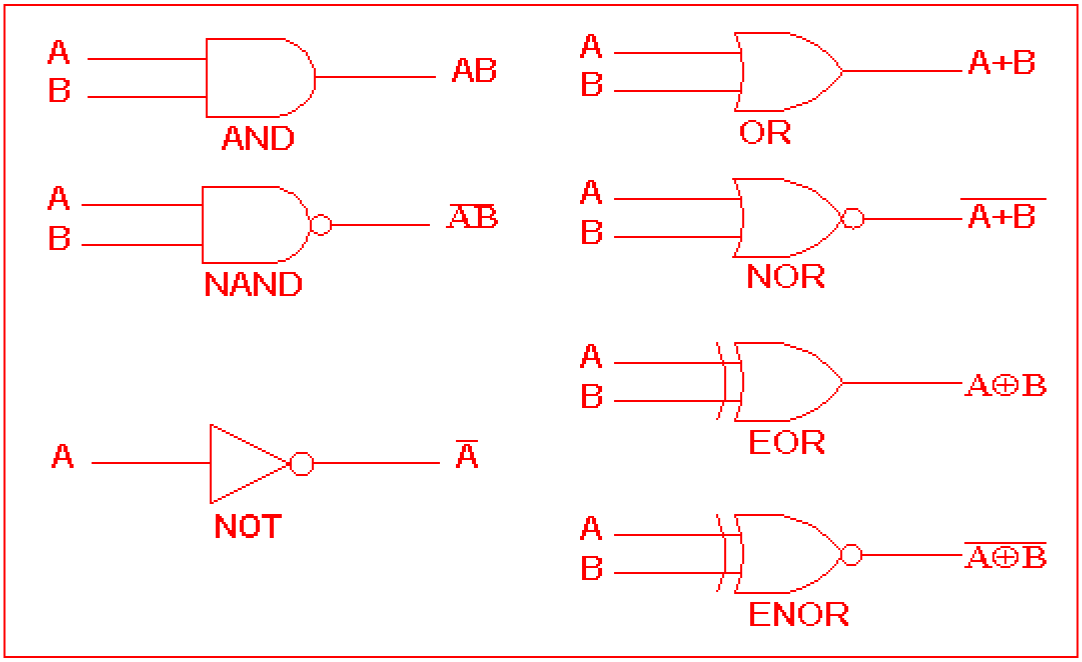
This is a NOT-AND gate which is equal to an AND gate followed by a NOT gate. The outputs of all NAND gates are high if **any** of the inputs are low. The symbol is an AND gate with a small circle on the output. The small circle represents inversion.

NOR gate



2 Input NOR gate		
A	B	$\overline{A+B}$
0	0	1
0	1	0
1	0	0
1	1	0

This is a NOT-OR gate which is equal to an OR gate followed by a NOT gate. The outputs of all NOR gates are low if **any** of the inputs are high. The symbol is an OR gate with a small circle on the output. The small circle represents inversion.



Logic gates representation using the Truth table

NOT gate	
A	$\overline{A}$
0	1
1	0

INPUTS		OUTPUTS					
A	B	AND	NAND	OR	NOR	EXOR	EXNOR
0	0	0	1	0	1	0	1
0	1	0	1	1	0	1	0
1	0	0	1	1	0	1	0
1	1	1	0	1	0	0	1

Integrated Circuit [IC]

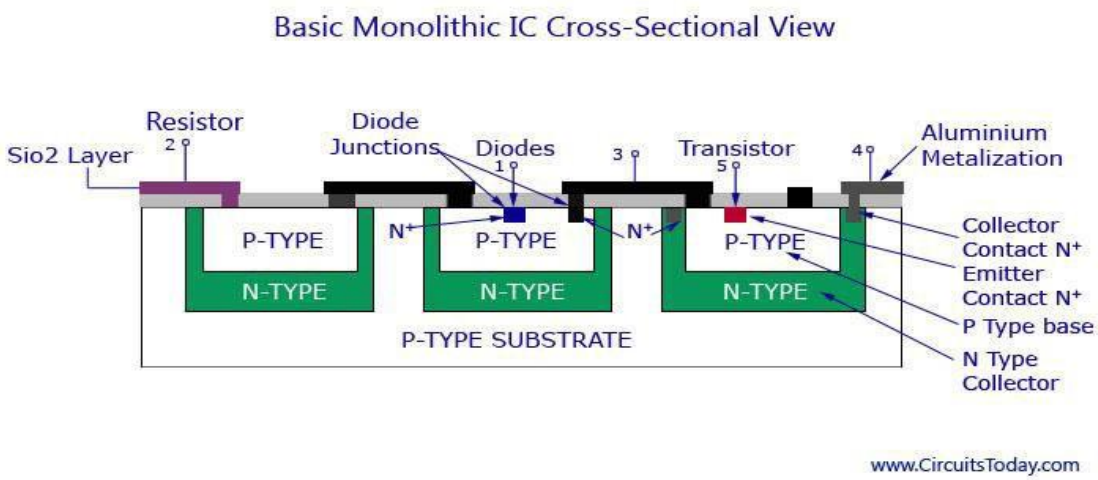
An integrated circuit (IC) is a small semiconductor-based electronic device consisting of fabricated transistors, resistors and capacitors. Integrated circuits are the building blocks of most electronic devices and equipment. An integrated circuit is also known as a chip or microchip.

An **Integrated Circuit (IC)** is an electronic device that gathers (or integrates) a few electronic components on a small semiconductor chip. Usually, an IC has a particular functionality. This functionality could be as specific as amplifying the voltage of a signal or applying a logic AND on 3 inputs and it could be broad as a microprocessor.

ICs can be digital or analog. Usually, analog ICs handle continuous signals, such as audio signals. Digital ICs handle discrete signals such as binary values.

Logic functions, such as AND, OR and NOT, are essential in building functionality for modern digital systems. So, how do you implement a logic function? Logic functions are implemented using transistors. For example, Figure 1 shows an internal transistor structure of a NAND gate.

Fabrication of Monolithic Integrated Circuits



A monolithic IC is one that has all components of a circuit and their interconnections made on a single wafer. They are the cheapest and thus more popular than other types of integrated circuits. The basic process of production of monolithic ICs is as follows.

1.P-substrate

The first step in the manufacture of monolithic ICs is called the p-substrate. In this step a wafer or substrate which is the bottom most layer of the circuit is made. A wafer is a polished silicon crystal cut in to required size. A single base or substrate can be cut and used tomake hundreds of ICs making monolithic ICs cheaper than other types. This Layer produces a high resistivity p-type semiconductor.

2.**EpitaxialGrowth**

In this step, the wafers are placed in a diffusion furnace of about 1,200oC. The furnace has a gas mixture containing silicon and pentavalent atoms which is passed over the wafers. The atoms form a thin layer of p-type semiconductor. All passive and active components of the circuit are fabricated. This layer is called the epitaxial layer.

3**. Insulating Layer**

The insulating layer is composed of silicon dioxide formed over the surface of the epitaxial layer to prevent the epitaxial layer from contamination. Pure oxygen is passed over the epitaxial layer to form the insulating layer.

5. **Component Diffusion**

Diffusion is the process of fabricating the circuit components on the wafer. The etched wafer is now sub-divided into portions, ones that cannot take impurities because they still have the SiO2 layer remaining. The second are diffusion acceptor portions that have the SiO2 removed.Controlled and very high temperatures are used to diffuse p-type impurities into the exposed n-type epitaxial layer. This result in n-type regions called isolation islands; they permit isolation of the various components of the IC, the components will later be formed separately on different islands. The diffusion process is repeated a number of times as various components are fabricated on to the IC. Aluminium is used to create the required interconnections between the components.

6. **Scribing and mounting**

The ready wafers now need to be put in chips. Hundreds of ICs can be produced from a single wafer. The wafer is then cut and assembled in the required chips. The chips have a ceramic casing and external leads are made after mounting the IC on to the cover, a process called encapsulating. The IC chips are now ready for use.

Advantages and Disadvantages Of Integrated Circuits

Here are some of the advantages observed in integrated circuits;

1. Since the soldering joints are not used in integrated circuits, this means that they are more reliable than discrete circuits. This is due to the reduction in number of interconnections between components.
2. Due to fabrication of the various components on the integrated circuits, the components became much smaller. This makes integrated circuits much lighter than discrete circuits. The integrated circuits thus consume much less space than discrete circuits.
3. Integrated circuits are encapsulated with a silicon oxide layer during manufacture. This layer is tough and resistant and thus gives the integrated an ability to operate at extremes of temperatures and other extreme environmental conditions.
4. Integrated circuits are constrained to minimize the number of external connections. This has greatly simplified the layout of these circuits and makes them easier to use.
5. Integrated circuits have been noticed to use less power for operations.
6. Integrated circuits have a lower cost of production than that of discrete circuits.

Disadvantages that occur in integrated circuits

1. If one component in an integrated fail, that means the whole circuit has to be replaced.

2. Integrated circuits have limited capacitances. This calls for external components if the capacitance needs an extension.

3. It is impossible to fabricate transformers or any other kind of inductor onto the integrated circuits and again calling for a discrete circuit.

4. Power that integrated circuits can produce is limited and calls for extension.

5. Integrated circuits are not flexible. Their components cannot be modified, and neither can the parameters of operation.

SCALE OF INTEGRATION

Small-scale integration (SSI)

The first integrated circuits contained only a few transistors. Early digital circuits containing tens of transistors provided a few logic gates, and early linear ICs such as the Plessey SL201 or the Philips TAA320 had as few as two transistors. The number of transistors in an integrated circuit has increased dramatically since then.

Medium-scale integration (MSI)

The next step in the development of integrated circuits introduced devices which contained hundreds of transistors on each chip, called "medium-scale integration" (MSI). MOSFET scaling technology made it possible to build high-density chips

Large-scale integration (LSI)

Further development, driven by the same MOSFET scaling technology and economic factors, led to "large-scale integration" (LSI) by the mid-1970s, with tens of thousands of transistors per chip. The masks used to process and manufacture SSI, MSI and early LSI and VLSI devices

Review Questions
Part A

1. Write a note on conventional energy?
2. Write a basic principle of solar drier and solar cell.
3. Explain the process of generation of bio gas?
4. What the basic principle of Optical fiber communication. Explain with a diagram.
5. Differentiate between the single mode fiber and multi-mode fiber.
6. Sketch the optic communication system block diagram. List out the parts.
7. Draw a diagram for NAND and NOR gate with Truth Table.
8. What is LED? What are the applications?
9. List out the advantages of IC
10. What you mean by breakdown voltage?

Part-B

1. Sketch the structure of solar water heater and explain the working of solar heater. What are the other applications?
2. Describe the process of wave propagation through single mode and multi-mode fibers. Sketch suitable diagram. Write advantages and disadvantages.
3. Draw a structure of zener diode. What are the input and out characteristics. Describe its function.