NEET UG 2023 H3 Question Paper with Solutions

Time Allowed :3 Hour 20 Minutes	Maximum Marks :720	Total Questions :200

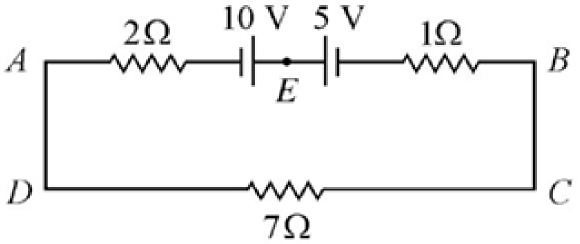
General Instructions

Read the following instructions very carefully and strictly follow them:

1. The Answer Sheet is this Test Booklet. When you are directed to open the Test Booklet, take the Answer Sheet and fill in the particulars in ORIGINAL Copy carefully with blue/black ball pen only.
2. The test is of 3 hours 20 minutes duration and the Test Booklet contains 200 multiple-choice questions (four options with a single correct answer) from Physics, Chemistry, and Biology (Botany and Zoology). 50 questions in each subject are divided into two Sections (A and B) as per details given below:
3. (a) Section A shall consist of 35 (Thirty-five) questions in each subject (Question Nos. 1 to 35, 51 to 85, 101 to 135 and 151 to 185).
4. (b) Section B shall consist of 15 (Fifteen) questions in each subject (Question Nos. 36 to 50, 86 to 100, 136 to 150 and 186 to 200). In Section B, a candidate needs to attempt any 10 (Ten) questions out of 15 (Fifteen) in each subject.
5. Candidates are advised to read all 15 questions in each subject of Section B before they start attempting the question paper. In the event of a candidate attempting more than ten questions, the first ten questions answered by the candidate shall be evaluated.
6. Each question carries 4 marks. For each correct response, the candidate will get 4 marks. For each incorrect response, one mark will be deducted from the total scores. The maximum marks are 720.
7. Rough work is to be done in the space provided for this purpose in the Test Booklet only.
8. On completion of the test, the candidate must hand over the Answer Sheet (ORIGINAL and OFFICE Copy) to the Invigilator before leaving the Room/Hall. The candidates are allowed to take away this Test Booklet with them.
9. Use of Electronic/Manual Calculator is prohibited.

Physics

1. The magnitude and direction of the current I in the following circuit is



(A) $\frac{5}{9}$ A from A to B through E
(B) 1.5 A from B to A through E
(C) 0.2 A from B to A through E
(D) 0.5 A from A to B through E

Correct Answer: (D) 0.5 A from A to B through E

Solution:

Step 1: Understanding the Question:
The problem asks for the magnitude and direction of the electric current flowing through the branch AEB of the given electrical circuit.

Step 2: Key Formula or Approach:
We can solve this circuit by applying Kirchhoff's Voltage Law (KVL). For a single loop, the simplified formula is $I = \frac{\text{Net EMF}}{\text{Total Resistance}}$. The net electromotive force (EMF) is the algebraic sum of the voltages of the batteries, and the total resistance is the sum of all resistances in the series circuit.

Step 3: Detailed Explanation:
The given circuit can be treated as a single series loop A-E-B-C-D-A.
The two voltage sources (batteries) are connected in opposition. The 10 V battery tries to push current in the clockwise direction (A to E to B), while the 5 V battery tries to push current in the counter-clockwise direction (B to E to A).
The net EMF in the circuit is the difference between the two battery voltages:

$$E_{net} = 10\text{ V} - 5\text{ V} = 5\text{ V}$$

Since the 10 V battery is stronger, the current will flow in the direction it dictates, which is clockwise (from A to B through E).
The resistors are all in series. The total resistance in the circuit is the sum of the individual resistances:

$$R_{total} = 2\,\Omega + 1\,\Omega + 7\,\Omega = 10\,\Omega$$

Now, we can calculate the magnitude of the current using Ohm's law for the entire circuit:

$$I = \frac{E_{net}}{R_{total}} = \frac{5\text{ V}}{10\,\Omega} = 0.5\text{ A}$$

Step 4: Final Answer:
The magnitude of the current is 0.5 A, and its direction is from A to B through E. This corresponds to option (D).

> **Quick Tip**
>
> In a simple single-loop circuit with multiple batteries, first find the net EMF by considering their polarities. If they support each other, add the EMFs; if they oppose, subtract them. The direction of the current will be dictated by the battery with the larger EMF.

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2. The net magnetic flux through any closed surface is :
(A) Infinity
(B) Negative
(C) Zero
(D) Positive

Correct Answer: (C) Zero

Solution:

Step 1: Understanding the Question:
This is a fundamental theoretical question about magnetic fields and their properties, specifically asking about the magnetic flux through a closed surface.

Step 2: Key Formula or Approach:
The concept is described by Gauss's Law for Magnetism, which is one of Maxwell's equations. The law is mathematically stated as:

$$\Phi_B = \oint \vec{B} \cdot d\vec{A} = 0$$

Where Φ_B is the magnetic flux, $\vec{B}$ is the magnetic field, and the integral is over a closed surface A.

Step 3: Detailed Explanation:
Gauss's Law for Magnetism states that the net magnetic flux out of any closed surface is zero. This is a consequence of the experimental observation that magnetic monopoles (isolated north or south poles) do not exist.
Magnetic field lines are always continuous loops; they do not begin or end at any point.
Therefore, for any closed surface, the number of magnetic field lines entering the surface is always equal to the number of magnetic field lines leaving it.
This results in a net magnetic flux of zero.

Step 4: Final Answer:
The net magnetic flux through any closed surface is always zero. This corresponds to option (C).

Quick Tip

Remember the contrast with Gauss's Law for electricity: the net electric flux through a closed surface is proportional to the enclosed electric charge ($\oint \vec{E} \cdot d\vec{A} = Q_{enc}/\epsilon_0$). Since there are no magnetic "charges" (monopoles), the magnetic equivalent is always zero.

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3. The amount of energy required to form a soap bubble of radius 2 cm from a soap solution is nearly: (surface tension of soap solution = 0.03 N m¹)
(A) 3.01×10^{-4} J
(B) 50.1×10^{-4} J
(C) 30.16×10^{-4} J
(D) 5.06×10^{-4} J

Correct Answer: (A) 3.01×10^{-4} J

Solution:

Step 1: Understanding the Question:
The question asks for the energy (or work done) needed to create a soap bubble of a given radius. This energy is stored as surface potential energy, which depends on the surface tension of the solution and the surface area of the bubble.

Step 2: Key Formula or Approach:
The energy required is equal to the work done against surface tension, given by the formula:

$$W = T \times \Delta A$$

where T is the surface tension and ΔA is the increase in the surface area.
Crucially, a soap bubble has two surfaces (an inner and an outer surface) in contact with air. Therefore, the total surface area is $A = 2 \times (4\pi r^2)$.

Step 3: Detailed Explanation:
First, convert the given radius from centimeters to meters:

$$r = 2 \text{ cm} = 2 \times 10^{-2} \text{ m}$$

Next, calculate the total surface area (ΔA) of the soap bubble. Since the bubble is formed from a solution (negligible initial area), the change in area is the final area of the bubble.

$$\Delta A = 2 \times (4\pi r^2) = 8\pi r^2$$

Substitute the value of the radius:

$$\Delta A = 8\pi(2 \times 10^{-2})^2 = 8\pi(4 \times 10^{-4}) = 32\pi \times 10^{-4} \text{ m}^2$$

Now, use the formula for the energy required, with the given surface tension $T = 0.03$ N m^{-1}:

$$W = T \times \Delta A = 0.03 \times (32\pi \times 10^{-4})$$

Using the approximation $\pi \approx 3.14$:

$$W = 0.03 \times 32 \times 3.14 \times 10^{-4} = 0.96 \times 3.14 \times 10^{-4} \approx 3.0144 \times 10^{-4} \text{ J}$$

Step 4: Final Answer:
The energy required is approximately 3.01×10^{-4} J. This corresponds to option (A).

Quick Tip

A common mistake is to forget that a soap bubble has two surfaces (inner and outer). A liquid drop in air has only one surface. Always double the area calculation for a bubble.

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4. A 12 V, 60 W lamp is connected to the secondary of a step down transformer, whose primary is connected to ac mains of 220 V. Assuming the transformer to be ideal, what is the current in the primary winding?
(A) 3.7 A
(B) 0.37 A
(C) 0.27 A
(D) 2.7 A

Correct Answer: (C) 0.27 A

Solution:

Step 1: Understanding the Question:
The problem asks for the current in the primary coil of an ideal transformer, given the input (primary) voltage and the output (secondary) power specifications.

Step 2: Key Formula or Approach:
For an ideal transformer, the efficiency is 100%, which means the power input to the primary coil is equal to the power output from the secondary coil.

$$P_{primary} = P_{secondary}$$

The power in the primary coil is given by:

$$P_{primary} = V_{primary} \times I_{primary}$$

Step 3: Detailed Explanation:
The lamp connected to the secondary coil has a power rating of 60 W. This is the output power from the transformer.

$$P_{secondary} = 60 \text{ W}$$

For an ideal transformer, the input power must equal the output power:

$$P_{primary} = P_{secondary} = 60 \text{ W}$$

The primary coil is connected to AC mains with a voltage of 220 V.

$$V_{primary} = 220 \text{ V}$$

Now, we can find the current in the primary winding using the power formula:

$$I_{primary} = \frac{P_{primary}}{V_{primary}}$$

$$I_{primary} = \frac{60 \text{ W}}{220 \text{ V}} = \frac{6}{22} = \frac{3}{11} \text{ A}$$

Converting the fraction to a decimal:

$$I_{primary} \approx 0.2727... \text{ A}$$

Step 4: Final Answer:
The current in the primary winding is approximately 0.27 A. This corresponds to option (C).

Quick Tip

For ideal transformer problems, the conservation of power ($P_{in} = P_{out}$ or $V_p I_p = V_s I_s$) is the key principle. You don't always need to calculate the turns ratio unless specifically asked.

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5. In a series LCR circuit, the inductance L is 10 mH, capacitance C is 1 μF and resistance R is 100 Ω. The frequency at which resonance occurs is:
(A) 1.59 rad/s
(B) 1.59 kHz
(C) 15.9 rad/s
(D) 15.9 kHz

Correct Answer: (B) 1.59 kHz

Solution:

Step 1: Understanding the Question:
The question asks for the resonant frequency of a series LCR circuit with given values for inductance (L), capacitance (C), and resistance (R).

Step 2: Key Formula or Approach:
Resonance in a series LCR circuit occurs when the inductive reactance (X_L) equals the capacitive reactance (X_C). The frequency at which this happens is the resonant frequency, f_0.
The formula for the resonant angular frequency is $\omega_0 = \frac{1}{\sqrt{LC}}$.
The resonant frequency is related to the angular frequency by $f_0 = \frac{\omega_0}{2\pi}$.
Therefore, the formula for resonant frequency is:

$$f_0 = \frac{1}{2\pi\sqrt{LC}}$$

Note that the resistance (R) does not affect the resonant frequency itself, but it does affect the sharpness of the resonance (the Q-factor).

Step 3: Detailed Explanation:
First, convert the given values of L and C to SI units.

$$L = 10 \text{ mH} = 10 \times 10^{-3} \text{ H}$$
$$C = 1 \text{ µF} = 1 \times 10^{-6} \text{ F}$$

Next, calculate the product LC:

$$LC = (10 \times 10^{-3}) \times (1 \times 10^{-6}) = 10 \times 10^{-9} = 10^{-8} \text{ s}^2$$

Now, calculate the square root of LC:

$$\sqrt{LC} = \sqrt{10^{-8}} = 10^{-4} \text{ s}$$

Finally, calculate the resonant frequency f_0:

$$f_0 = \frac{1}{2\pi\sqrt{LC}} = \frac{1}{2\pi \times 10^{-4}} = \frac{10^4}{2\pi} \text{ Hz}$$

Using the approximation $\pi \approx 3.14$:

$$f_0 = \frac{10000}{2 \times 3.14} = \frac{10000}{6.28} \approx 1592.3 \text{ Hz}$$

To express this in kilohertz (kHz), divide by 1000:

$$f_0 \approx 1.5923 \text{ kHz}$$

Step 4: Final Answer:
The resonant frequency is approximately 1.59 kHz. This corresponds to option (B). Note that options (A) and (C) are in rad/s, which represent angular frequency (ω), not frequency (f).

Quick Tip

Always be mindful of the units in the options. Questions may mix frequency (in Hz) and angular frequency (in rad/s). Remember the conversion $\omega = 2\pi f$. Also, ensure all your input values (L, C) are in base SI units (Henrys, Farads) before calculating.

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6. Given below are two statements:
Statement I: Photovoltaic devices can convert optical radiation into electricity.
Statement II: Zener diode is designed to operate under reverse bias in breakdown region.
In the light of the above statements, choose the most appropriate answer from the options given below :
(A) Statement I is correct but Statement II is incorrect.
(B) Statement I is incorrect but Statement II is correct.
(C) Both Statement I and Statement II are correct.
(D) Both Statement I and Statement II are incorrect.

Correct Answer: (C) Both Statement I and Statement II are correct.

Solution:

Step 1: Understanding the Question:
The question presents two statements related to semiconductor devices and asks to evaluate their correctness.

Step 2: Detailed Explanation:
Analysis of Statement I:
"Photovoltaic devices can convert optical radiation into electricity."
This statement describes the fundamental principle of the photovoltaic effect. Devices like solar cells are specifically designed p-n junctions that, when exposed to light (optical radiation), generate a voltage and can produce an electric current. Thus, they convert light energy directly into electrical energy. Statement I is correct.

Analysis of Statement II:
"Zener diode is designed to operate under reverse bias in breakdown region."
A Zener diode is a special type of diode. While a standard diode is not meant to operate in the reverse breakdown region (as it can be destructive), a Zener diode is specifically manufactured to have a precise and stable reverse breakdown voltage (the Zener voltage). Its primary application, such as in voltage regulation, relies on its ability to conduct current in the reverse direction when the voltage across it reaches the Zener voltage, thus operating in the breakdown region without being damaged. Statement II is correct.

Step 3: Final Answer:
Since both Statement I and Statement II are correct descriptions of their respective devices,

the correct option is (C).

Quick Tip

For statement-based questions in physics, focus on the core definition and primary application of the concepts or devices mentioned. Photovoltaic = light to electricity. Zener diode = reverse breakdown voltage regulation.

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7. The temperature of a gas is -50° C. To what temperature the gas should be heated so that the rms speed is increased by 3 times?
(A) 3097 K
(B) 223 K
(C) 669° C
(D) 3295° C

Correct Answer: (D) 3295° C

Solution:

Step 1: Understanding the Question:
The question asks for the final temperature required to increase the root-mean-square (rms) speed of gas molecules. The increase is specified as "increased by 3 times."

Step 2: Key Formula or Approach:
The rms speed (v_{rms}) of gas molecules is related to the absolute temperature (T) by the formula:

$$v_{rms} = \sqrt{\frac{3RT}{M}}$$

From this, we can see the proportionality:

$$v_{rms} \propto \sqrt{T}$$

where T must be in Kelvin. This gives the relationship $\frac{v_1}{v_2} = \sqrt{\frac{T_1}{T_2}}$.

Step 3: Detailed Explanation:
First, convert the initial temperature from Celsius to Kelvin.

$$T_1 = -50^\circ\text{C} + 273.15 \approx 223 \text{ K}$$

The phrase "increased by 3 times" means the final speed is the initial speed plus three times the initial speed.

$$v_2 = v_1 + 3v_1 = 4v_1$$

So, the ratio of the final speed to the initial speed is $\frac{v_2}{v_1} = 4$.
Now use the proportionality relationship:

$$\left(\frac{v_2}{v_1}\right)^2 = \frac{T_2}{T_1}$$

Substitute the known values:

$$(4)^2 = \frac{T_2}{223 \text{ K}}$$
$$16 = \frac{T_2}{223 \text{ K}}$$

Solve for the final temperature in Kelvin, T_2:

$$T_2 = 16 \times 223 \text{ K} = 3568 \text{ K}$$

The options are given in both Kelvin and Celsius. Let's convert T_2 to Celsius.

$$T_2(^\circ\text{C}) = 3568 - 273.15 \approx 3295^\circ\text{C}$$

Step 4: Final Answer:
The final temperature required is 3568 K or approximately 3295° C. This matches option (D).

Quick Tip

Always convert temperatures to Kelvin for calculations involving gas laws and kinetic theory. Also, be very careful with phrases like "increased by X times" (new = old + X*old) versus "increased to X times" (new = X*old). Here, the former interpretation fits the provided answer.

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8. The venturi-meter works on :
(A) The principle of parallel axes
(B) The principle of perpendicular axes
(C) Huygen's principle
(D) Bernoulli's principle

Correct Answer: (D) Bernoulli's principle

Solution:

Step 1: Understanding the Question:
This is a direct question asking for the physical principle behind the operation of a venturi-meter.

Step 2: Detailed Explanation:
A venturi-meter is a device used to measure the rate of flow of a fluid through a pipe. It consists

of a tube with a constricted section (the throat).
As the fluid passes through the throat, its velocity increases due to the principle of continuity (A1v1 = A2v2).
According to **Bernoulli's principle**, for a horizontal pipe, where the fluid velocity is higher, the pressure is lower. The principle states:

$$P + \frac{1}{2}\rho v^2 + \rho gh = \text{constant}$$

By measuring the pressure difference between the wider part of the tube and the throat, one can calculate the fluid's velocity and thus its flow rate.
The other options are unrelated:
- The principles of parallel and perpendicular axes are theorems used to calculate the moment of inertia of rigid bodies in mechanics.
- Huygen's principle is a concept in optics that describes how waves propagate.

Step 3: Final Answer:
The operation of a venturi-meter is based on Bernoulli's principle. This corresponds to option (D).

Quick Tip

Associate common physics instruments with their underlying principles. For example: Venturi-meter -¿ Bernoulli's Principle; Spectrometer -¿ Dispersion; Potentiometer -¿ Null Deflection Method.

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9. A vehicle travels half the distance with speed v and the remaining distance with speed 2v. Its average speed is:
(A) $\frac{40}{3}$
(B) $\frac{30}{4}$
(C) $\frac{v}{3}$
(D) $\frac{20}{3}$

Correct Answer: (A) $\frac{40}{3}$

Solution:

Step 1: Understanding the Question:
The problem asks for the average speed of a vehicle that covers two equal distances at two different speeds, given as v and $2v$.

Step 2: Key Formula or Approach:
When an object travels two equal distances with speeds v_1 and v_2, the average speed is the harmonic mean of the individual speeds.
Let the total distance be $2d$. The first half of the distance (d) is covered with speed v_1, and

the second half (d) with speed v_2.
Time taken for the first half, $t_1 = \frac{d}{v_1}$.
Time taken for the second half, $t_2 = \frac{d}{v_2}$.
Average speed $v_{avg} = \frac{\text{Total Distance}}{\text{Total Time}} = \frac{2d}{t_1+t_2} = \frac{2d}{\frac{d}{v_1}+\frac{d}{v_2}} = \frac{2}{\frac{1}{v_1}+\frac{1}{v_2}} = \frac{2v_1v_2}{v_1+v_2}$.

Step 3: Detailed Explanation:
In this specific problem, we are given $v_1 = v$ and $v_2 = 2v$.
Substituting these values into the average speed formula:

$$v_{avg} = \frac{2(v)(2v)}{v+2v} = \frac{4v^2}{3v} = \frac{4v}{3}$$

The calculated average speed is $\frac{4v}{3}$, which is a symbolic expression depending on v. However, the options provided are numerical, and the answer key points to $\frac{40}{3}$ as the correct answer. This indicates a likely typo or missing information in the question as presented. The question is symbolic, but the answer is numerical.
For the answer to be $\frac{40}{3}$, we can equate our derived expression to the given answer:

$$\frac{4v}{3} = \frac{40}{3}$$
$$4v = 40$$
$$v = 10$$

This implies that the question likely intended for the speeds to be $v = 10$ units (e.g., m/s or km/h) and $2v = 20$ units. Let's verify with these numerical values.
If $v_1 = 10$ and $v_2 = 20$, then:

$$v_{avg} = \frac{2 \times 10 \times 20}{10+20} = \frac{400}{30} = \frac{40}{3}$$

This matches the correct answer.

Step 4: Final Answer:
Assuming the implicit value of $v = 10$, the speeds are 10 and 20. The average speed is $\frac{40}{3}$. This corresponds to option (A).

Quick Tip

Be aware of potential typos in exam questions. If your derived symbolic answer (e.g., $\frac{4v}{3}$) doesn't match symbolic options, but a numerical option matches if you assume a simple value for the variable (like v=10), it's a strong hint about the question's intent. When distances are equal, use the harmonic mean formula for average speed.

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10. An ac source is connected to a capacitor C. Due to decrease in its operating frequency :

(A) displacement current decreases.
(B) capacitive reactance remains constant
(C) capacitive reactance decreases.
(D) displacement current increases.

Correct Answer: (A) displacement current decreases.

Solution:

Step 1: Understanding the Question:
The question asks how the displacement current and capacitive reactance of a capacitor change when the frequency of the connected AC source decreases.

Step 2: Key Formula or Approach:
The capacitive reactance (X_C) is given by the formula:

$$X_C = \frac{1}{\omega C} = \frac{1}{2\pi f C}$$

where f is the frequency and C is the capacitance.
The current (I) in a purely capacitive circuit is given by Ohm's law:

$$I = \frac{V}{X_C}$$

The displacement current (I_d) inside the capacitor is equal to the conduction current (I) in the connecting wires.

$$I_d = I$$

Step 3: Detailed Explanation:
1. **Effect on Capacitive Reactance:** From the formula $X_C = \frac{1}{2\pi f C}$, it is clear that capacitive reactance is inversely proportional to the frequency ($X_C \propto \frac{1}{f}$).
If the operating frequency f decreases, the capacitive reactance X_C will increase. This contradicts options (B) and (C).

2. **Effect on Current:** The current flowing through the circuit is $I = \frac{V}{X_C} = V(2\pi f C)$.
This shows that the current I is directly proportional to the frequency f.
Therefore, as the frequency f decreases, the current I in the circuit also decreases.

3. **Effect on Displacement Current:** The displacement current I_d through the capacitor is equal to the conduction current I flowing in the wires.
Since the conduction current I decreases, the displacement current I_d also decreases.

Step 4: Final Answer:
A decrease in operating frequency leads to an increase in capacitive reactance, which in turn causes a decrease in the current. Since the displacement current is equal to the conduction current, the displacement current decreases. Thus, option (A) is correct.

Quick Tip

Remember the relationships for capacitors and inductors in AC circuits. For a capacitor, $X_C \propto 1/f$, so at low frequencies, it acts like an open circuit (high reactance). For an inductor, $X_L \propto f$, so at low frequencies, it acts like a short circuit (low reactance).

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11. Light travels a distance x in time t_1 in air and 10x in time t in another denser medium. What is the critical angle for this medium?

(A) $\sin^{-1}\left(\frac{t_1}{10t_2}\right)$
(B) $\sin^{-1}\left(\frac{10t_1}{t_2}\right)$
(C) $\sin^{-1}\left(\frac{t_2}{t_1}\right)$
(D) $\sin^{-1}\left(\frac{10t_2}{t_1}\right)$

Correct Answer: (B) $\sin^{-1}\left(\frac{10t_1}{t_2}\right)$

Solution:

Step 1: Understanding the Question:
The question asks for the critical angle of a denser medium, given the distance and time taken by light to travel in air (rarer medium) and in the denser medium. The critical angle is the angle of incidence in the denser medium for which the angle of refraction in the rarer medium is 90 degrees.

Step 2: Key Formula or Approach:
1. Speed of light in a medium: $v = \frac{\text{distance}}{\text{time}}$.
2. Refractive index of a medium: $n = \frac{\text{speed of light in vacuum (c)}}{\text{speed of light in medium (v)}}$.
3. The relative refractive index of medium 2 (denser) with respect to medium 1 (rarer) is $n_{21} = \frac{n_2}{n_1} = \frac{v_1}{v_2}$.
4. The formula for the critical angle (C) is $\sin(C) = \frac{n_{rarer}}{n_{denser}} = \frac{n_1}{n_2}$.

Step 3: Detailed Explanation:
First, calculate the speed of light in both media.
Speed in air (medium 1): $v_1 = \frac{x}{t_1}$. The refractive index of air is $n_1 \approx 1$.
Speed in the denser medium (medium 2): $v_2 = \frac{10x}{t_2}$.

Next, find the refractive index of the denser medium with respect to air (n_2/n_1).

$$\frac{n_2}{n_1} = \frac{v_1}{v_2} = \frac{x/t_1}{10x/t_2} = \frac{x}{t_1} \times \frac{t_2}{10x} = \frac{t_2}{10t_1}$$

Now, use the formula for the critical angle.

$$\sin(C) = \frac{n_{rarer}}{n_{denser}} = \frac{n_1}{n_2}$$

This is the reciprocal of the value we just calculated.

$$\sin(C) = \frac{1}{n_2/n_1} = \frac{1}{t_2/(10t_1)} = \frac{10t_1}{t_2}$$

Finally, find the critical angle C by taking the inverse sine.

$$C = \sin^{-1}\left(\frac{10t_1}{t_2}\right)$$

Step 4: Final Answer:

The critical angle for the medium is $\sin^{-1}\left(\frac{10t_1}{t_2}\right)$, which corresponds to option (B).

Quick Tip

For critical angle problems, always remember that $\sin(C) = \frac{n_{rarer}}{n_{denser}}$. The value of $\sin(C)$ must be less than 1, which means $n_{rarer} < n_{denser}$. This can be a quick check for your calculated refractive index ratio.

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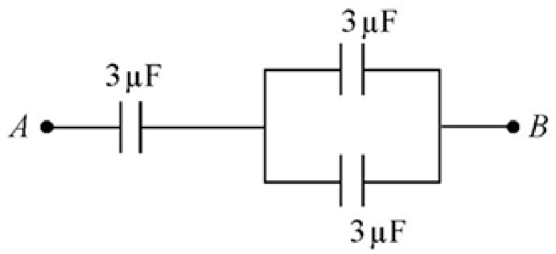
12. The equivalent capacitance of the system shown in the following circuit is :



(A) 6 µF
(B) 9 µF
(C) 2 µF
(D) 3 µF

Correct Answer: (C) 2 µF

Solution:

Step 1: Understanding the Question:

The question asks for the equivalent capacitance between points A and B for the given circuit configuration.

Step 2: Key Formula or Approach:

1. For capacitors in parallel, the equivalent capacitance is the sum of individual capacitances: $C_p = C_1 + C_2 + \ldots$.

2. For capacitors in series, the reciprocal of the equivalent capacitance is the sum of the reciprocals of individual capacitances: $\frac{1}{C_s} = \frac{1}{C_1} + \frac{1}{C_2} + \dots$.

Step 3: Detailed Explanation:
Let's analyze the circuit diagram. Let the junction point after the first capacitor be P.
The circuit has a 3 µF capacitor between A and P.
From point P, the circuit splits into two parallel branches that both connect to point B. Each branch contains a 3 µF capacitor.
Let's call the first capacitor $C_1 = 3\,\mu F$, the top capacitor in the parallel branch $C_2 = 3\,\mu F$, and the bottom capacitor in the parallel branch $C_3 = 3\,\mu F$.

First, calculate the equivalent capacitance of the parallel combination of C_2 and C_3.

$$C_{23} = C_2 + C_3 = 3\,\mu F + 3\,\mu F = 6\,\mu F$$

Now, this equivalent capacitance C_{23} is in series with the first capacitor C_1.
The total equivalent capacitance C_{eq} between A and B is found by the series formula.

$$\frac{1}{C_{eq}} = \frac{1}{C_1} + \frac{1}{C_{23}}$$

$$\frac{1}{C_{eq}} = \frac{1}{3\,\mu F} + \frac{1}{6\,\mu F}$$

To add these fractions, we find a common denominator, which is 6.

$$\frac{1}{C_{eq}} = \frac{2}{6\,\mu F} + \frac{1}{6\,\mu F} = \frac{3}{6\,\mu F} = \frac{1}{2\,\mu F}$$

Therefore, the equivalent capacitance is:

$$C_{eq} = 2\,\mu F$$

Step 4: Final Answer:
The equivalent capacitance of the system is 2 µF. This corresponds to option (C).

Quick Tip

Remember that the rules for combining capacitors are opposite to those for resistors. Capacitors in parallel add up directly ($C_p = C_1 + C_2$), while for resistors in parallel, you add reciprocals. Capacitors in series add by reciprocals ($1/C_s = 1/C_1 + 1/C_2$), while resistors in series add directly.

---

13. The magnetic energy stored in an inductor of inductance 4 µH carrying a current of 2 A is :

(A) 8 mJ
(B) 8 µJ

(C) 4 µJ
(D) 4 mJ

Correct Answer: (B) 8 µJ

Solution:

Step 1: Understanding the Question:
The question asks for the amount of magnetic potential energy stored in an inductor with a given inductance and current.

Step 2: Key Formula or Approach:
The energy (U) stored in the magnetic field of an inductor is given by the formula:

$$U = \frac{1}{2}LI^2$$

where L is the inductance and I is the current flowing through it.

Step 3: Detailed Explanation:
We are given the following values:
Inductance, $L = 4\,\mu H = 4 \times 10^{-6}\,H$.
Current, $I = 2\,A$.

Substitute these values into the energy formula:

$$U = \frac{1}{2} \times (4 \times 10^{-6}\,H) \times (2\,A)^2$$

$$U = \frac{1}{2} \times (4 \times 10^{-6}) \times 4$$

$$U = 2 \times 10^{-6} \times 4$$

$$U = 8 \times 10^{-6}\,J$$

Since $1\,\mu J = 10^{-6}\,J$, the energy stored is:

$$U = 8\,\mu J$$

Step 4: Final Answer:
The magnetic energy stored in the inductor is 8 µJ. This corresponds to option (B).

Quick Tip

Pay close attention to the units and prefixes (µ for micro $= 10^{-6}$, m for milli $= 10^{-3}$). A common mistake is to misinterpret the prefix, leading to an answer that is off by a factor of 1000. In this case, confusing µJ and mJ would lead to choosing the wrong option.

---

14. A full wave rectifier circuit consists of two p-n junction diodes, a centre-tapped transformer, capacitor and a load resistance. Which of these components remove the ac ripple from the rectified output?

(A) Capacitor
(B) Load resistance
(C) A centre-tapped transformer
(D) p-n junction diodes

Correct Answer: (A) Capacitor

Solution:

Step 1: Understanding the Question:
The question asks to identify the component in a full-wave rectifier circuit that is responsible for filtering, i.e., removing the AC component (ripple) from the pulsating DC output.

Step 2: Detailed Explanation:
Let's analyze the role of each component in a full-wave rectifier circuit:
1. **A centre-tapped transformer:** This component steps down the AC mains voltage to a suitable level and provides two AC signals that are 180 degrees out of phase, which are necessary for the two diodes to conduct in alternate half-cycles. It does not filter the output.
2. **p-n junction diodes:** These diodes are the core of the rectifier. They allow current to flow in only one direction, converting the AC input into a pulsating DC output. They perform the rectification but do not remove the ripple.
3. **Load resistance (R_L):** This is the resistance across which the output voltage is obtained. It does not have a filtering function.
4. **Capacitor:** A capacitor connected in parallel with the load resistance acts as a filter. It gets charged to the peak voltage when the rectified voltage is increasing. When the rectified voltage starts to decrease, the capacitor begins to discharge slowly through the load resistance. This process fills in the "valleys" between the peaks of the pulsating DC, significantly reducing the AC ripple and smoothing the output voltage to a more constant DC level. This is often called a smoothing capacitor or filter capacitor.

Step 3: Final Answer:
The component that removes the AC ripple from the rectified output is the capacitor. Thus, option (A) is correct.

Quick Tip

In electronics, capacitors are often used to block DC while passing AC, or to filter/smooth varying DC voltages. Inductors can also be used in filter circuits (choke filters), as they oppose changes in current, which also helps in smoothing the output. The most common simple filter is the capacitor filter described here.

15. In a plane electromagnetic wave travelling in free space, the electric field component oscillates sinusoidally at a frequency of 2.0×10^{10} Hz and amplitude 48 Vm^{-1}. Then the amplitude of oscillating magnetic field is : (Speed of light in free space $= 3 \times 10^8$ m s^{-1})

(A) 1.6×10^{-7} T
(B) 1.6×10^{-6} T
(C) 1.6×10^{-9} T
(D) 1.6×10^{-8} T

Correct Answer: (A) 1.6×10^{-7} T

Solution:

Step 1: Understanding the Question:
The question provides the amplitude of the electric field component of an electromagnetic wave in free space and asks for the amplitude of the magnetic field component.

Step 2: Key Formula or Approach:
In an electromagnetic wave propagating in a vacuum (free space), the ratio of the amplitudes of the electric field (E_0) and the magnetic field (B_0) is equal to the speed of light (c).

$$c = \frac{E_0}{B_0}$$

Step 3: Detailed Explanation:
We are given:
Amplitude of the electric field, $E_0 = 48\,\text{V/m}$.
Speed of light in free space, $c = 3 \times 10^8\,\text{m/s}$.
The frequency of oscillation (2.0×10^{10} Hz) is extra information not needed for this calculation.

We need to find the amplitude of the magnetic field, B_0. Rearranging the formula:

$$B_0 = \frac{E_0}{c}$$

Substitute the given values:

$$B_0 = \frac{48}{3 \times 10^8}$$

$$B_0 = 16 \times 10^{-8}\,\text{T}$$

To express this in standard scientific notation, we adjust the decimal point:

$$B_0 = 1.6 \times 10^1 \times 10^{-8}\,\text{T}$$

$$B_0 = 1.6 \times 10^{-7}\,\text{T}$$

Step 4: Final Answer:
The amplitude of the oscillating magnetic field is 1.6×10^{-7} T. This corresponds to option (A).

Quick Tip

A simple way to remember the relationship is E=cB (often remembered as "Electromagnetic waves Can be Boring"). This helps to avoid inverting the ratio E_0/B_0 to B_0/E_0. Also, remember that the numerical value of E_0 is always much larger than B_0 because c is a very large number.

---

16. The errors in the measurement which arise due to unpredictable fluctuations in temperature and voltage supply are :

(A) Least count errors
(B) Random errors
(C) Instrumental errors
(D) Personal errors

Correct Answer: (B) Random errors

Solution:

Step 1: Understanding the Question:
The question asks to classify the type of measurement error that results from unpredictable changes in environmental or experimental conditions, such as temperature and voltage.

Step 2: Detailed Explanation:
Let's define the different types of errors listed in the options:
1. **Least count errors:** This error is associated with the resolution of the instrument. The least count is the smallest value that can be measured by the measuring instrument. This type of error is systematic in nature.
2. **Instrumental errors:** These errors arise from imperfections in the design or calibration of the measuring instrument. For example, a zero error in a vernier caliper. These are also a type of systematic error.
3. **Personal errors:** These errors are introduced due to the fault of the observer, such as parallax error (improper eye positioning), bias, or carelessness in taking readings.
4. **Random errors:** These are errors that occur irregularly and are random in magnitude and sign. They are caused by sudden and unpredictable fluctuations in experimental conditions like temperature, pressure, voltage supply, or mechanical vibrations. Because they are unpredictable, they cannot be eliminated but can be minimized by taking multiple observations and calculating their mean.

The question specifically mentions "unpredictable fluctuations in temperature and voltage supply," which perfectly matches the definition of random errors.

Step 3: Final Answer:
Errors arising from unpredictable fluctuations in experimental conditions are classified as random errors. Therefore, option (B) is correct.

Quick Tip

A key distinction between systematic errors and random errors is that systematic errors are consistent and repeatable (e.g., always too high or always too low), while random errors fluctuate and are equally likely to be positive or negative. Taking the average of many measurements reduces random errors but does not affect systematic errors.

---

17. Let a wire be suspended from the ceiling (rigid support) and stretched by a weight W attached at its free end. The longitudinal stress at any point of cross-sectional area A of the wire is :

(A) W/2A
(B) Zero
(C) 2W/A
(D) W/A

Correct Answer: (D) W/A

Solution:

Step 1: Understanding the Question:
The question asks for the formula for longitudinal stress in a wire that is stretched by a suspended weight.

Step 2: Key Formula or Approach:
Longitudinal stress (σ) is defined as the restoring force (F) acting per unit cross-sectional area (A) of the material.

$$\sigma = \frac{\text{Force}}{\text{Area}} = \frac{F}{A}$$

Step 3: Detailed Explanation:
1. A wire is suspended vertically, and a weight W is attached to its free end. This weight acts as the external deforming force, pulling the wire downwards.
2. The wire is in equilibrium, which means at any cross-section, an internal restoring force is developed that is equal in magnitude and opposite in direction to the external deforming force.
3. The external deforming force is the weight W. Therefore, the internal restoring force is $F = W$.
4. The cross-sectional area of the wire is given as A.

5. Substituting the force and area into the stress formula:

$$\sigma = \frac{F}{A} = \frac{W}{A}$$

This stress is called tensile stress, which is a type of longitudinal stress.

Step 4: Final Answer:
The longitudinal stress at any point of the wire is W/A. This corresponds to option (D).

Quick Tip

Stress and Pressure have the same dimensions ($[ML^{-1}T^{-2}]$) and units (Pascals or N/m²). However, stress is a tensor quantity that describes internal forces in a material, while pressure is a scalar quantity, typically used for fluids. In this simple case, the magnitude is calculated similarly.

---

18. Resistance of a carbon resistor determined from colour codes is (22000±5%) Ω. The colour of third band must be :

(A) Orange
(B) Yellow
(C) Red
(D) Green

Correct Answer: (A) Orange

Solution:

Step 1: Understanding the Question:
The question provides the value of a carbon resistor and asks to identify the color of its third band based on the standard resistor color code system.

Step 2: Key Formula or Approach:
The value of a four-band resistor is given by: (First band digit)(Second band digit) $\times 10^{\text{Third band}}$ $\Omega \pm$ (Fourth band tolerance %).
We need to recall the color code chart for digits and multipliers.

- Black: 0
- Brown: 1
- Red: 2
- **Orange: 3**

- Yellow: 4
- Green: 5
- Blue: 6
- Violet: 7
- Grey: 8
- White: 9

Step 3: Detailed Explanation:
The given resistance is $22000\,\Omega \pm 5\%$.
Let's break down the resistance value $22000\,\Omega$ into the color code format.

$$22000 = 22 \times 1000 = 22 \times 10^3$$

1. **First Band:** The first significant digit is 2. The color for the digit 2 is **Red**.
2. **Second Band:** The second significant digit is 2. The color for the digit 2 is **Red**.
3. **Third Band (Multiplier):** The multiplier is 10^3. The color corresponding to a multiplier of 10^3 is **Orange**.
4. **Fourth Band (Tolerance):** The tolerance is $\pm 5\%$. The color for a 5% tolerance is **Gold**.

The question specifically asks for the color of the third band. The third band represents the multiplier 10^3, which corresponds to the color Orange.

Step 4: Final Answer:
The color of the third band must be Orange. This corresponds to option (A).

Quick Tip

A popular mnemonic to remember the resistor color code is: "**B**ig **B**oys **R**ace **O**ur **Y**oung **G**irls **B**ut **V**iolet **G**enerally **W**ins". The first letter of each word corresponds to the colors Black, Brown, Red, Orange, Yellow, Green, Blue, Violet, Grey, White for digits 0 through 9.

---

19. The work functions of Caesium (Cs), Potassium (K) and Sodium (Na) are 2.14 eV, 2.30 eV and 2.75 eV respectively. If incident electromagnetic radiation has an incident energy of 2.20 eV, which of these photosensitive surfaces may emit photoelectrons?
(A) K only
(B) Na only
(C) Cs only
(D) Both Na and K

Correct Answer: (C) Cs only

Solution:

Step 1: Understanding the Question:
The question is about the photoelectric effect. We need to determine which of the given metals (Caesium, Potassium, Sodium) will emit photoelectrons when illuminated by radiation of a specific energy.

Step 2: Key Formula or Approach:
The condition for the photoelectric effect to occur is that the energy of the incident radiation (E) must be greater than or equal to the work function (Φ) of the metal surface.

$$E \geq \Phi$$

The work function is the minimum energy required to remove an electron from the surface of a material.

Step 3: Detailed Explanation:
The energy of the incident radiation is given as $E = 2.20$ eV.
The work functions for the metals are:
- Caesium (Cs): $\Phi_{Cs} = 2.14$ eV
- Potassium (K): $\Phi_K = 2.30$ eV
- Sodium (Na): $\Phi_{Na} = 2.75$ eV

Now, we check the condition $E > \Phi$ for each metal:
- For Caesium (Cs): 2.20 eV > 2.14 eV. The condition is satisfied. Photoelectrons will be emitted.
- For Potassium (K): 2.20 eV < 2.30 eV. The condition is not satisfied. No photoelectrons will be emitted.
- For Sodium (Na): 2.20 eV < 2.75 eV. The condition is not satisfied. No photoelectrons will be emitted.

Step 4: Final Answer:
Only Caesium (Cs) will emit photoelectrons. This corresponds to option (C).

Quick Tip

To solve photoelectric effect problems, simply remember the threshold condition: Incident Energy must be greater than the Work Function. Any excess energy becomes the kinetic energy of the emitted electron.

---

20. For Young's double slit experiment, two statements are given below:
Statement I: If screen is moved away from the plane of slits, angular separation of the fringes remains constant.
Statement II: If the monochromatic source is replaced by another monochromatic source of larger wavelength, the angular separation of fringes decreases.

In the light of the above statements, choose the correct answer from the options given below:
(A) Statement I is true but Statement II is false.
(B) Statement I is false but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (A) Statement I is true but Statement II is false.

Solution:

Step 1: Understanding the Question:
The question asks to evaluate two statements concerning the angular separation of fringes in a Young's double-slit experiment (YDSE).

Step 2: Key Formula or Approach:
In YDSE, the fringe width is given by $\beta = \frac{\lambda D}{d}$.
The angular separation (or angular fringe width) is given by $\theta = \frac{\beta}{D}$.
Substituting the expression for β, we get:

$$\theta = \frac{(\lambda D/d)}{D} = \frac{\lambda}{d}$$

where λ is the wavelength of light and d is the distance between the slits.

Step 3: Detailed Explanation:
Analysis of Statement I:
"If screen is moved away from the plane of slits, angular separation of the fringes remains constant."
The distance of the screen from the slits is D. The formula for angular separation is $\theta = \frac{\lambda}{d}$. This formula does not depend on D. Therefore, moving the screen away (changing D) does not change the angular separation. Statement I is true.

Analysis of Statement II:
"If the monochromatic source is replaced by another monochromatic source of larger wavelength, the angular separation of fringes decreases."
The angular separation is $\theta = \frac{\lambda}{d}$. This shows that the angular separation θ is directly proportional to the wavelength λ. If the wavelength is made larger, the angular separation will increase, not decrease. Therefore, Statement II is false.

Step 4: Final Answer:
Statement I is true and Statement II is false. This corresponds to option (A).

Quick Tip

Distinguish carefully between fringe width (β) and angular fringe width (θ). Fringe width (β) depends on the screen distance D, but angular width (θ) does not. Both are directly proportional to the wavelength λ.

---

21. The half life of a radioactive substance is 20 minutes. In how much time, the activity of substance drops to $(\frac{1}{16})^{th}$ of its initial value?

(A) 60 minutes
(B) 80 minutes
(C) 20 minutes
(D) 40 minutes

Correct Answer: (B) 80 minutes

Solution:

Step 1: Understanding the Question:
The question asks for the total time elapsed for the activity of a radioactive substance to decrease to 1/16 of its initial value, given its half-life.

Step 2: Key Formula or Approach:
The relationship between the final activity (A), initial activity (A), elapsed time (t), and half-life (T/) is given by the formula:

$$\frac{A}{A_0} = \left(\frac{1}{2}\right)^{t/T_{1/2}}$$

Alternatively, we can express the fraction remaining as $(1/2)^n$, where n is the number of half-lives that have passed. Then, $t = n \times T_{1/2}$.

Step 3: Detailed Explanation:
We are given that the final activity is 1/16 of the initial activity.

$$\frac{A}{A_0} = \frac{1}{16}$$

We can write 1/16 as a power of 1/2:

$$\frac{1}{16} = \frac{1}{2^4} = \left(\frac{1}{2}\right)^4$$

Comparing this with the decay formula, we see that the number of half-lives, n, is 4.

$$n = \frac{t}{T_{1/2}} = 4$$

The half-life $T_{1/2}$ is given as 20 minutes.
The total time t can be calculated as:

$$t = n \times T_{1/2} = 4 \times 20 \text{ minutes} = 80 \text{ minutes}$$

Step 4: Final Answer:
The time required for the activity to drop to 1/16 of its initial value is 80 minutes. This corresponds to option (B).

Quick Tip

For fractions that are integer powers of 2 (like 1/2, 1/4, 1/8, 1/16, etc.), it's quickest to count the number of half-lives. $1/16 = (1/2)^4$, so 4 half-lives have passed.

---

22. A football player is moving southward and suddenly turns eastward with the same speed to avoid an opponent. The force that acts on the player while turning is :
(A) along north-east
(B) along south-west
(C) along eastward
(D) along northward

Correct Answer: (A) along north-east

Solution:

Step 1: Understanding the Question:
The question asks for the direction of the net force acting on a player during a change in direction, which implies a change in velocity (acceleration).

Step 2: Key Formula or Approach:
According to Newton's second law, the net force ($\vec{F}$) acting on an object is proportional to its acceleration ($\vec{a}$), and the force is in the same direction as the acceleration.

$$\vec{F} = m\vec{a}$$

Acceleration is the rate of change of velocity: $\vec{a} = \frac{\Delta \vec{v}}{\Delta t}$.
Therefore, the direction of the force is the same as the direction of the change in velocity, $\Delta \vec{v} = \vec{v}_{final} - \vec{v}_{initial}$.

Step 3: Detailed Explanation:
Let's set up a coordinate system. Let the eastward direction be the positive x-axis ($+\hat{i}$) and the northward direction be the positive y-axis ($+\hat{j}$).
- The player is initially moving southward. So, the initial velocity is $\vec{v}_{initial} = -v\hat{j}$, where v is the speed.
- The player then turns to move eastward with the same speed. So, the final velocity is $\vec{v}_{final} = +v\hat{i}$.

Now, calculate the change in velocity $\Delta\vec{v}$:

$$\Delta\vec{v} = \vec{v}_{final} - \vec{v}_{initial} = (v\hat{i}) - (-v\hat{j}) = v\hat{i} + v\hat{j}$$

The direction of the force is the direction of $\Delta\vec{v}$. The vector $v\hat{i} + v\hat{j}$ has a positive x-component (east) and a positive y-component (north). A vector with equal positive east and north components points in the north-east direction.

Step 4: Final Answer:
The force acts along the north-east direction. This corresponds to option (A).

Quick Tip

For problems involving change in velocity, always perform vector subtraction ($\vec{v}_{final} - \vec{v}_{initial}$). A quick way to visualize this is to draw the vectors: draw $\vec{v}_{final}$ and then add $-\vec{v}_{initial}$ to its tail. The resultant vector gives the direction of the change.

---

23. In hydrogen spectrum, the shortest wavelength in the Balmer series is λ. The shortest wavelength in the Bracket series is :
(A) 9λ
(B) 16λ
(C) 2λ
(D) 4λ

Correct Answer: (D) 4λ

Solution:

Step 1: Understanding the Question:
The question relates the shortest wavelength of the Balmer series to the shortest wavelength of the Brackett series in the hydrogen spectrum.

Step 2: Key Formula or Approach:
The Rydberg formula for the wavelength of spectral lines in hydrogen is:

$$\frac{1}{\lambda} = R\left(\frac{1}{n_1^2} - \frac{1}{n_2^2}\right)$$

where R is the Rydberg constant, and n_1 and n_2 are integers with $n_2 > n_1$.
- For the Balmer series, $n_1 = 2$. - For the Brackett series, $n_1 = 4$. The shortest wavelength (λ_{min}) in any series occurs for the largest energy transition, which corresponds to the electron coming from $n_2 = \infty$. This is called the series limit.

Step 3: Detailed Explanation:
For the Balmer series:

The shortest wavelength ($\lambda_{B,min}$) is when $n_1 = 2$ and $n_2 = \infty$.

$$\frac{1}{\lambda_{B,min}} = R\left(\frac{1}{2^2} - \frac{1}{\infty^2}\right) = R\left(\frac{1}{4} - 0\right) = \frac{R}{4}$$

The problem states this wavelength is λ. So, $\lambda = \frac{4}{R}$.

For the Brackett series:
The shortest wavelength ($\lambda_{Br,min}$) is when $n_1 = 4$ and $n_2 = \infty$.

$$\frac{1}{\lambda_{Br,min}} = R\left(\frac{1}{4^2} - \frac{1}{\infty^2}\right) = R\left(\frac{1}{16} - 0\right) = \frac{R}{16}$$

So, $\lambda_{Br,min} = \frac{16}{R}$.

Relating the two wavelengths:
We want to express $\lambda_{Br,min}$ in terms of λ.

$$\lambda_{Br,min} = \frac{16}{R} = 4 \times \left(\frac{4}{R}\right) = 4\lambda$$

Step 4: Final Answer:
The shortest wavelength in the Brackett series is 4λ. This corresponds to option (D).

Quick Tip

Remember the first level (n_1) for the main spectral series: Lyman (1), Balmer (2), Paschen (3), Brackett (4), Pfund (5). The shortest wavelength (highest energy) for any series is the "series limit," found by setting $n_2 = \infty$.

---

24. Two bodies of mass m and 9m are placed at a distance R. The gravitational potential on the line joining the bodies where the gravitational field equals zero, will be (G = gravitational constant) :
(A) $-\frac{16Gm}{R}$
(B) $-\frac{20Gm}{R}$
(C) $-\frac{8Gm}{R}$
(D) $-\frac{12Gm}{R}$

Correct Answer: (A) $-\frac{16Gm}{R}$

Solution:

Step 1: Understanding the Question:
The problem has two parts. First, find the point on the line joining two masses (m and 9m) where the net gravitational field is zero. Second, calculate the total gravitational potential at

that point.

Step 2: Key Formula or Approach:
- Gravitational field due to a point mass M at distance r is $E = \frac{GM}{r^2}$ (magnitude). The field is a vector. - Gravitational potential due to a point mass M at distance r is $V = -\frac{GM}{r}$. Potential is a scalar.

Step 3: Detailed Explanation:
Part 1: Find the point of zero gravitational field.
Let mass m be at position x=0 and mass 9m be at x=R. The point of zero field must be between the two masses. Let this point be at a distance r from mass m. Its distance from mass 9m will be $R - r$.
The gravitational fields from the two masses are in opposite directions at this point. For the net field to be zero, their magnitudes must be equal.

$$E_m = E_{9m}$$
$$\frac{Gm}{r^2} = \frac{G(9m)}{(R-r)^2}$$
$$\frac{1}{r^2} = \frac{9}{(R-r)^2}$$

Taking the square root of both sides:

$$\frac{1}{r} = \frac{3}{R-r}$$
$$R - r = 3r$$
$$R = 4r \implies r = \frac{R}{4}$$

So, the point is at a distance of R/4 from mass m and R - R/4 = 3R/4 from mass 9m.

Part 2: Calculate the potential at this point.
The total potential is the scalar sum of the potentials from each mass.

$$V_{total} = V_m + V_{9m}$$
$$V_{total} = \left(-\frac{Gm}{r}\right) + \left(-\frac{G(9m)}{R-r}\right)$$

Substitute the distances we found: $r = R/4$ and $R - r = 3R/4$.

$$V_{total} = \left(-\frac{Gm}{R/4}\right) + \left(-\frac{9Gm}{3R/4}\right)$$
$$V_{total} = -\frac{4Gm}{R} - \frac{36Gm}{3R}$$
$$V_{total} = -\frac{4Gm}{R} - \frac{12Gm}{R}$$
$$V_{total} = -\frac{16Gm}{R}$$

Step 4: Final Answer:
The gravitational potential at the point where the field is zero is $-\frac{16Gm}{R}$. This corresponds to option (A).

Quick Tip

For two masses, the point of zero gravitational field is always closer to the smaller mass. Remember that gravitational field is a vector (magnitudes are equated for the null point), while potential is a scalar (values are added algebraically).

---

25. The minimum wavelength of X-rays produced by an electron accelerated through a potential difference of V volts is proportional to:
(A) $\frac{1}{\sqrt{V}}$
(B) V^2
(C) $\sqrt{V}$
(D) $\frac{1}{V}$

Correct Answer: (D) $\frac{1}{V}$

Solution:

Step 1: Understanding the Question:
The question asks for the relationship between the minimum wavelength of produced X-rays and the accelerating potential difference applied to the electrons.

Step 2: Key Formula or Approach:
When an electron is accelerated through a potential difference V, it gains kinetic energy (KE) given by:

$$KE = eV$$

where e is the charge of the electron.
When this electron strikes a target, it decelerates, and its kinetic energy is converted into electromagnetic radiation (X-rays). The most energetic X-ray photon is produced when the electron loses all its kinetic energy in a single collision. This corresponds to the minimum wavelength (λ_{min}) of the X-rays, also known as the cutoff wavelength.
The energy of a photon is given by $E = hf = \frac{hc}{\lambda}$.
By conservation of energy:

$$KE = E_{photon,max}$$
$$eV = \frac{hc}{\lambda_{min}}$$

Step 3: Detailed Explanation:
From the energy conservation equation, we can express the minimum wavelength λ_{min} in terms

of the accelerating voltage V:

$$\lambda_{min} = \frac{hc}{eV}$$

Here, h (Planck's constant), c (speed of light), and e (electron charge) are all constants. Therefore, we can write the proportionality:

$$\lambda_{min} \propto \frac{1}{V}$$

The minimum wavelength of the X-rays is inversely proportional to the accelerating potential difference.

Step 4: Final Answer:
The minimum wavelength of X-rays is proportional to $\frac{1}{V}$. This corresponds to option (D).

Quick Tip

This relationship, $\lambda_{min} = \frac{hc}{eV}$, is fundamental to the continuous X-ray spectrum. A useful value to remember for quick calculations is $hc \approx 1240$ eV · nm. Then, $\lambda_{min}(\text{in nm}) = \frac{1240}{V(\text{in volts})}$.

---

26. The ratio of radius of gyration of a solid sphere of mass M and radius R about its own axis to the radius of gyration of the thin hollow sphere of same mass and radius about its axis is :
(A) 2:5
(B) 5:2
(C) 3:5
(D) 5:3
(E)$\sqrt{3} : \sqrt{5}$

Correct Answer: (E) $\sqrt{3} : \sqrt{5}$

Solution:

Step 1: Understanding the Question:
The question asks for the ratio of the radius of gyration of a solid sphere to that of a hollow sphere, both having the same mass M and radius R, and rotating about their own axes (diameters).

Step 2: Key Formula or Approach:
The radius of gyration (k) is related to the moment of inertia (I) and mass (M) by the formula $I = Mk^2$, which means $k = \sqrt{\frac{I}{M}}$.
- Moment of inertia of a solid sphere about its axis: $I_{solid} = \frac{2}{5}MR^2$
- Moment of inertia of a thin hollow sphere (spherical shell) about its axis: $I_{hollow} = \frac{2}{3}MR^2$

Step 3: Detailed Explanation:
First, find the radius of gyration for the solid sphere (k_{solid}):

$$k_{solid}^2 = \frac{I_{solid}}{M} = \frac{\frac{2}{5}MR^2}{M} = \frac{2}{5}R^2$$
$$k_{solid} = \sqrt{\frac{2}{5}}R$$

Next, find the radius of gyration for the hollow sphere (k_{hollow}):

$$k_{hollow}^2 = \frac{I_{hollow}}{M} = \frac{\frac{2}{3}MR^2}{M} = \frac{2}{3}R^2$$
$$k_{hollow} = \sqrt{\frac{2}{3}}R$$

Now, find the required ratio of the radii of gyration:

$$\frac{k_{solid}}{k_{hollow}} = \frac{\sqrt{\frac{2}{5}}R}{\sqrt{\frac{2}{3}}R} = \sqrt{\frac{2/5}{2/3}} = \sqrt{\frac{2}{5} \times \frac{3}{2}} = \sqrt{\frac{3}{5}}$$

The exact ratio is $\sqrt{3} : \sqrt{5}$.

Quick Tip

Be prepared for occasional errors or ambiguities in exam questions. If your correct derivation leads to an answer not in the options, re-read the question and see if a slightly different interpretation (like ratio of I instead of k, or k^2) matches an option. This is a common exam-taking strategy.

---

27. A metal wire has mass (0.4 $\pm$ 0.002) g, radius (0.3$\pm$0.001) mm and length (5$\pm$0.02) cm. The maximum possible percentage error in the measurement of density will nearly be:

(A) 1.6%
(B) 1.4%
(C) 1.2%
(D) 1.3%

Correct Answer: (A) 1.6%

Solution:

Step 1: Understanding the Question:
The question requires us to calculate the maximum possible percentage error in the density of a wire, given the values and errors for its mass, radius, and length.

Step 2: Key Formula or Approach:
The density (ρ) of a cylindrical wire is given by the formula $\rho = \frac{\text{mass}}{\text{volume}} = \frac{m}{V}$.
The volume of the wire is $V = \pi r^2 l$, where r is the radius and l is the length.
So, the formula for density is $\rho = \frac{m}{\pi r^2 l}$.
The formula for the maximum relative error in density is given by the sum of the relative errors of the individual quantities, with each error multiplied by the power to which the quantity is raised:

$$\frac{\Delta\rho}{\rho} = \frac{\Delta m}{m} + 2\frac{\Delta r}{r} + \frac{\Delta l}{l}$$

To find the percentage error, we multiply the relative error by 100.

$$\%\text{ error in }\rho = \left(\frac{\Delta m}{m} + 2\frac{\Delta r}{r} + \frac{\Delta l}{l}\right) \times 100\%$$

Step 3: Detailed Explanation:
First, we identify the given values:
Mass, $m = 0.4$ g, with error $\Delta m = 0.002$ g.
Radius, $r = 0.3$ mm, with error $\Delta r = 0.001$ mm.
Length, $l = 5$ cm, with error $\Delta l = 0.02$ cm.

Next, we calculate the individual percentage errors for each measurement:
Percentage error in mass:

$$\frac{\Delta m}{m} \times 100 = \frac{0.002}{0.4} \times 100 = 0.5\%$$

Percentage error in radius:

$$\frac{\Delta r}{r} \times 100 = \frac{0.001}{0.3} \times 100 = \frac{1}{3}\% \approx 0.333\%$$

Percentage error in length:

$$\frac{\Delta l}{l} \times 100 = \frac{0.02}{5} \times 100 = 0.4\%$$

Now, we substitute these values into the formula for the percentage error in density:

$$\%\text{ error in }\rho = \left(\frac{\Delta m}{m} \times 100\right) + 2\left(\frac{\Delta r}{r} \times 100\right) + \left(\frac{\Delta l}{l} \times 100\right)$$

$$\%\text{ error in }\rho = 0.5\% + 2(0.333\%) + 0.4\%$$

$$\%\text{ error in }\rho = 0.5\% + 0.666\% + 0.4\%$$

$$\%\text{ error in }\rho = 1.566\%$$

Step 4: Final Answer:
The calculated percentage error is approximately 1.566%, which is nearly 1.6%.
Therefore, the correct option is (A).

Quick Tip

When calculating percentage error for a quantity derived from a formula, remember to add the percentage errors of the individual measurements. If a quantity is raised to a power 'n' (like radius squared in the volume formula), its percentage error is multiplied by 'n'.

---

28. If $\oint_S \vec{E} \cdot d\vec{S} = 0$ over a surface, then :

(A) all the charges must necessarily be inside the surface.
(B) the electric field inside the surface is necessarily uniform.
(C) the number of flux lines entering the surface must be equal to the number of flux lines leaving it.
(D) the magnitude of electric field on the surface is constant.

Correct Answer: (C) the number of flux lines entering the surface must be equal to the number of flux lines leaving it.

Solution:

Step 1: Understanding the Question:
The question is based on Gauss's Law in electrostatics. The expression $\oint_S \vec{E} \cdot d\vec{S}$ represents the total electric flux through a closed surface S. We are given that this total flux is zero and need to determine the correct conclusion from the given options.

Step 2: Key Formula or Approach:
Gauss's Law states that the total electric flux through any closed surface is proportional to the net electric charge enclosed by that surface:

$$\oint_S \vec{E} \cdot d\vec{S} = \frac{Q_{enc}}{\epsilon_0}$$

where Q_{enc} is the net charge inside the surface and ϵ_0 is the permittivity of free space.

Step 3: Detailed Explanation:
Given the condition $\oint_S \vec{E} \cdot d\vec{S} = 0$, from Gauss's Law, we can conclude that:

$$\frac{Q_{enc}}{\epsilon_0} = 0 \implies Q_{enc} = 0$$

This means the net charge enclosed by the surface is zero. Let's analyze the options based on this conclusion.

(A) This is incorrect. Zero net charge inside does not mean there are no charges. It could mean there's an equal amount of positive and negative charge (e.g., an electric dipole) inside the surface, or no charges at all. Also, charges can exist outside the surface.

(B) This is incorrect. The electric field inside is not necessarily uniform. For example, if an electric dipole is placed inside the surface, the net enclosed charge is zero, but the electric field inside is not uniform.

(C) This is the correct interpretation of zero net flux. Electric flux represents the number of electric field lines passing through a surface. A positive flux means more lines are leaving the surface than entering it, while a negative flux means more lines are entering than leaving. If the net flux is zero, it implies that the number of field lines entering the surface is exactly equal to the number of field lines leaving it.

(D) This is incorrect. The electric field on the surface is not necessarily constant. Consider a Gaussian surface with an external point charge nearby. The net flux is zero, but the field strength on the surface will be different at different points.

Step 4: Final Answer:
The condition of zero total electric flux means that the inward flux and outward flux are equal in magnitude. Therefore, the number of flux lines entering the surface must be equal to the number of flux lines leaving it.

Quick Tip

Remember that Gauss's law relates the flux through a closed surface to the *net* charge *inside* that surface. A net flux of zero implies a net enclosed charge of zero, which is visually represented as an equal number of electric field lines entering and leaving the surface.

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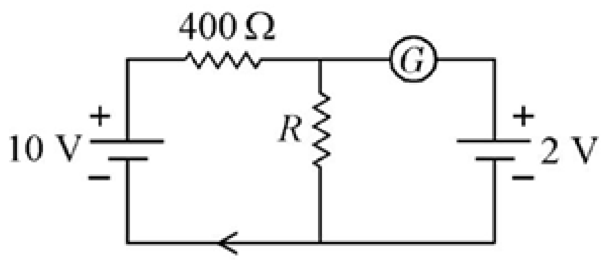
29. If the galvanometer G does not show any deflection in the circuit shown, the value of R is given by:



(A) 100 Ω
(B) 400 Ω
(C) 200 Ω
(D) 50 Ω

Correct Answer: (A) 100 Ω

Solution:

Step 1: Understanding the Question:
The problem describes a circuit and states that the galvanometer (G) shows no deflection. This is a key condition which implies that no current is flowing through the galvanometer. We need to find the value of the unknown resistance R based on this condition.

Step 2: Key Formula or Approach:
If there is no current flowing through the galvanometer, the potential at the two points connected by the galvanometer must be equal. This is the principle of a potentiometer or a balanced bridge. We can apply Kirchhoff's laws or simply equate the potentials at the two terminals of the galvanometer.

Step 3: Detailed Explanation:
Let's label the points in the circuit. Let the point between the 400 Ω resistor and the galvanometer be A. The point on the other side of the galvanometer is connected to the positive terminal of the 2 V battery. Let's call this point B.
The condition "galvanometer G does not show any deflection" means the current through G is zero. This implies that the potential at point A is equal to the potential at point B ($V_A = V_B$).

Let's assume the negative terminal of both batteries is at a potential of 0 V (ground).
The potential at point B is therefore the potential of the positive terminal of the 2 V battery, so $V_B = 2$ V.
Since $V_A = V_B$, we have $V_A = 2$ V.

Now let's analyze the left loop containing the 10 V battery, the 400 Ω resistor, and the resistor R. Since no current flows into the galvanometer branch, the current I flowing through the 400 Ω resistor is the same as the current flowing through R.
This current is given by Ohm's law for the entire left loop:

$$I = \frac{\text{Total Voltage}}{\text{Total Resistance}} = \frac{10 \text{ V}}{400\,\Omega + R}$$

The potential at point A can be calculated as the potential drop across the resistor R. The potential at the bottom wire is 0 V. So,

$$V_A = I \times R = \left(\frac{10}{400 + R}\right) R$$

We already established that $V_A = 2$ V. So, we can set up the equation:

$$2 = \frac{10R}{400 + R}$$

Now, we solve for R:

$$2(400 + R) = 10R$$

$$800 + 2R = 10R$$

$$800 = 10R - 2R$$

$$800 = 8R$$
$$R = \frac{800}{8} = 100\,\Omega$$

Step 4: Final Answer:
The value of the resistance R is 100 Ω.

Quick Tip

The phrase "no deflection in the galvanometer" is a strong hint. It immediately tells you that the potential difference across the galvanometer is zero. Use this to equate the potentials at the two connecting points and solve the problem. This is a more direct approach than setting up complex loop equations.

---

30. The potential energy of a long spring when stretched by 2 cm is U. If the spring is stretched by 8 cm, potential energy stored in it will be :

(A) 8U
(B) 16U
(C) 2U
(D) 4U

Correct Answer: (B) 16U

Solution:

Step 1: Understanding the Question:
The question asks to find the new potential energy stored in a spring when its stretch is increased, given its initial potential energy at a smaller stretch.

Step 2: Key Formula or Approach:
The potential energy (E_p) stored in a spring is given by the formula:

$$E_p = \frac{1}{2}kx^2$$

where k is the spring constant and x is the displacement (stretch or compression) from the equilibrium position.
From this formula, we can see that the potential energy is directly proportional to the square of the stretch: $E_p \propto x^2$.

Step 3: Detailed Explanation:
Let the initial state be denoted by subscript 1 and the final state by subscript 2.
Given:
Initial stretch, $x_1 = 2$ cm.
Initial potential energy, $E_{p1} = U$.

Final stretch, $x_2 = 8$ cm.

We need to find the final potential energy, E_{p2}.
Using the proportionality $E_p \propto x^2$, we can write a ratio:

$$\frac{E_{p2}}{E_{p1}} = \frac{\frac{1}{2}kx_2^2}{\frac{1}{2}kx_1^2} = \left(\frac{x_2}{x_1}\right)^2$$

Substitute the given values:

$$\frac{E_{p2}}{U} = \left(\frac{8 \text{ cm}}{2 \text{ cm}}\right)^2$$
$$\frac{E_{p2}}{U} = (4)^2 = 16$$
$$E_{p2} = 16U$$

Step 4: Final Answer:
The potential energy stored in the spring when stretched by 8 cm will be 16U.

Quick Tip

Recognize the quadratic relationship between spring potential energy and stretch ($U \propto x^2$). If the stretch is multiplied by a factor 'n', the energy is multiplied by n^2. Here, the stretch increases by a factor of 4 (from 2 cm to 8 cm), so the energy increases by a factor of $4^2 = 16$.

---

31. A Carnot engine has an efficiency of 50% when its source is at a temperature 327° C. The temperature of the sink is :

(A) 100° C
(B) 200° C
(C) 27° C
(D) 15° C

Correct Answer: (C) 27° C

Solution:

Step 1: Understanding the Question:
The question provides the efficiency and the source temperature of a Carnot engine and asks for the temperature of the sink. A crucial point in thermodynamics is to work with absolute temperatures (Kelvin).

Step 2: Key Formula or Approach:
The efficiency (η) of a Carnot engine is defined in terms of the source temperature (T_1) and

the sink temperature (T_2). The formula is:

$$\eta = 1 - \frac{T_2}{T_1}$$

where T_1 and T_2 must be in Kelvin.
The conversion from Celsius (°C) to Kelvin (K) is: $T(\text{K}) = T(°\text{C}) + 273$.

Step 3: Detailed Explanation:
First, let's list the given information and convert the units.
Efficiency, $\eta = 50\% = 0.50$.
Source temperature, $T_1 = 327°$ C.
Convert T_1 to Kelvin:

$$T_1(\text{K}) = 327 + 273 = 600 \text{ K}$$

Now, use the efficiency formula to find the sink temperature T_2 in Kelvin.

$$\eta = 1 - \frac{T_2}{T_1}$$
$$0.50 = 1 - \frac{T_2}{600}$$

Rearrange the formula to solve for T_2:

$$\frac{T_2}{600} = 1 - 0.50 = 0.50$$
$$T_2 = 0.50 \times 600 = 300 \text{ K}$$

The question asks for the sink temperature in degrees Celsius. So, we convert T_2 back to Celsius.

$$T_2(°\text{C}) = T_2(\text{K}) - 273$$
$$T_2(°\text{C}) = 300 - 273 = 27°\text{C}$$

Step 4: Final Answer:
The temperature of the sink is 27° C.

Quick Tip

Always convert temperatures to Kelvin when working with gas laws or thermodynamics formulas like the Carnot efficiency. This is one of the most common pitfalls in such problems. After finding the answer in Kelvin, remember to check if the question asks for the final answer in Celsius.

---

32. The angular acceleration of a body, moving along the circumference of a circle, is :

(A) along the tangent to its position
(B) along the axis of rotation

(C) along the radius, away from centre
(D) along the radius towards the centre

Correct Answer: (B) along the axis of rotation

Solution:

Step 1: Understanding the Question:
The question asks for the direction of the angular acceleration vector for a body in circular motion. It's important to distinguish between linear and angular quantities.

Step 2: Key Formula or Approach:
Angular velocity ($\vec{\omega}$) and angular acceleration ($\vec{\alpha}$) are axial vectors. Their direction is perpendicular to the plane of rotation and lies along the axis of rotation. The direction is determined by the right-hand rule.
Angular acceleration is defined as the rate of change of angular velocity: $\vec{\alpha} = \frac{d\vec{\omega}}{dt}$.

Step 3: Detailed Explanation:
For a body moving in a circle, the plane of motion is the plane of the circle. The axis of rotation is a line passing through the center of the circle and perpendicular to its plane.
- The angular velocity vector $\vec{\omega}$ always points along this axis of rotation.
- Since angular acceleration $\vec{\alpha}$ is the time derivative of angular velocity $\vec{\omega}$, its direction is also along the axis of rotation (provided the axis itself isn't changing direction).
- If the body is speeding up (angular speed increases), $\vec{\alpha}$ is in the same direction as $\vec{\omega}$.
- If the body is slowing down (angular speed decreases), $\vec{\alpha}$ is in the opposite direction to $\vec{\omega}$.
In either case, $\vec{\alpha}$ is directed along the axis of rotation.

Let's analyze the other options:
(A) along the tangent: This is the direction of linear tangential velocity ($\vec{v}$) and tangential acceleration ($\vec{a}_t$).
(C) & (D) along the radius: This is the direction of the position vector and the centripetal (or radial) acceleration ($\vec{a}_c$), which always points towards the center.

Step 4: Final Answer:
The angular acceleration of a body in circular motion is directed along the axis of rotation.

Quick Tip

Remember the distinction: linear kinematic quantities ($\vec{v}$, $\vec{a}$) lie in the plane of motion. Rotational or angular kinematic quantities ($\vec{\omega}$, $\vec{\alpha}$) are axial vectors and are perpendicular to the plane of motion, pointing along the axis of rotation.

---

33. The ratio of frequencies of fundamental harmonic produced by an open pipe to that of closed pipe having the same length is :

(A) 1:3
(B) 3:1
(C) 1:2
(D) 2:1

Correct Answer: (D) 2:1

Solution:

Step 1: Understanding the Question:
We need to find the ratio of the fundamental frequencies for two types of organ pipes (one open at both ends, one closed at one end) that have the same length.

Step 2: Key Formula or Approach:
The fundamental frequency corresponds to the longest possible wavelength (λ) that can form a standing wave in the pipe.
- **Open Pipe (open at both ends):** Antinodes form at both ends. The simplest standing wave pattern has one node in the middle. The length of the pipe L is half a wavelength: $L = \frac{\lambda_{open}}{2}$. The fundamental frequency is $f_{open} = \frac{v}{\lambda_{open}} = \frac{v}{2L}$.
- **Closed Pipe (closed at one end):** A node forms at the closed end and an antinode at the open end. The simplest standing wave pattern is a quarter of a wavelength: $L = \frac{\lambda_{closed}}{4}$. The fundamental frequency is $f_{closed} = \frac{v}{\lambda_{closed}} = \frac{v}{4L}$.
Here, v is the speed of sound.

Step 3: Detailed Explanation:
Let L be the common length of the pipes.
Fundamental frequency of the open pipe:

$$f_{open} = \frac{v}{2L}$$

Fundamental frequency of the closed pipe:

$$f_{closed} = \frac{v}{4L}$$

The question asks for the ratio of the open pipe's frequency to the closed pipe's frequency ($f_{open} : f_{closed}$).

$$\frac{f_{open}}{f_{closed}} = \frac{\left(\frac{v}{2L}\right)}{\left(\frac{v}{4L}\right)}$$

$$\frac{f_{open}}{f_{closed}} = \frac{v}{2L} \times \frac{4L}{v} = \frac{4}{2} = 2$$

So, the ratio is 2:1.

Step 4: Final Answer:
The ratio of the fundamental frequency of an open pipe to that of a closed pipe of the same

length is 2:1.

Quick Tip

A simple way to remember is: $f_{open} = 2 \times f_{closed}$ for the same length L. The fundamental frequency of an open pipe is twice that of a closed pipe. Also, remember that a closed pipe can only produce odd harmonics (1st, 3rd, 5th, ...), while an open pipe can produce all harmonics (1st, 2nd, 3rd, ...).

---

34. A bullet is fired from a gun at the speed of 280 m s^{-1} in the direction 30° above the horizontal. The maximum height attained by the bullet is (g = 9.8 m s^{-2}, sin 30° = 0.5) :

(A) 1000 m
(B) 3000 m
(C) 2800 m
(D) 2000 m

Correct Answer: (A) 1000 m

Solution:

Step 1: Understanding the Question:
This is a standard problem in projectile motion. We are given the initial velocity (speed and angle) of a projectile and asked to find the maximum vertical height it reaches.

Step 2: Key Formula or Approach:
The formula for the maximum height (H) reached by a projectile launched with initial speed u at an angle θ to the horizontal is:

$$H = \frac{u^2 \sin^2 \theta}{2g}$$

where g is the acceleration due to gravity.

Step 3: Detailed Explanation:
Let's identify the given values:
Initial speed, $u = 280$ m/s.
Angle of projection, $\theta = 30°$.
Value of $g = 9.8$ m/s^2.
We are also given that $\sin 30° = 0.5$.

Now, we substitute these values into the maximum height formula:

$$H = \frac{(280)^2 (\sin 30°)^2}{2 \times 9.8}$$

$$H = \frac{(280 \times 280) \times (0.5)^2}{19.6}$$
$$H = \frac{78400 \times 0.25}{19.6}$$
$$H = \frac{19600}{19.6}$$

To simplify the calculation:

$$H = \frac{196000}{196} = 1000 \text{ m}$$

Step 4: Final Answer:
The maximum height attained by the bullet is 1000 m.

> **Quick Tip**
>
> In projectile motion, the vertical motion determines the maximum height and time of flight, while the horizontal motion determines the range. The maximum height is reached when the vertical component of velocity becomes zero. The formula $H = \frac{u_y^2}{2g}$ where $u_y = u \sin\theta$ is very useful.

---

35. An electric dipole is placed at an angle of 30° with an electric field of intensity 2×10^5N C^{-1}. It experiences a torque equal to 4 Nm. Calculate the magnitude of charge on the dipole, if the dipole length is 2 cm.

(A) 4 mC
(B) 2 mC
(C) 8 mC
(D) 6 mC

Correct Answer: (B) 2 mC

Solution:

Step 1: Understanding the Question:
The question asks for the magnitude of the charge on an electric dipole, given the torque it experiences in a uniform electric field, the field strength, the angle, and the dipole length.

Step 2: Key Formula or Approach:
The torque (τ) on an electric dipole in an electric field (E) is given by:

$$\tau = pE \sin\theta$$

where p is the electric dipole moment and θ is the angle between the dipole moment and the electric field.

The electric dipole moment p is defined as:

$$p = q \times d$$

where q is the magnitude of the charge and d is the separation between the charges (dipole length).
Combining these, we get: $\tau = (qd)E\sin\theta$.

Step 3: Detailed Explanation:
First, list the given values and ensure they are in SI units.
Torque, $\tau = 4$ Nm.
Electric field, $E = 2 \times 10^5$ N/C.
Angle, $\theta = 30°$.
Dipole length, $d = 2$ cm $= 0.02$ m $= 2 \times 10^{-2}$ m.

We need to find the charge q. Let's rearrange the combined formula to solve for q:

$$q = \frac{\tau}{dE\sin\theta}$$

Substitute the given values into this equation:

$$q = \frac{4}{(2 \times 10^{-2}) \times (2 \times 10^5) \times \sin 30°}$$

We know that $\sin 30° = 0.5$.

$$q = \frac{4}{(2 \times 10^{-2}) \times (2 \times 10^5) \times 0.5}$$

$$q = \frac{4}{4 \times 10^3 \times 0.5}$$

$$q = \frac{4}{2 \times 10^3}$$

$$q = 2 \times 10^{-3} \text{ C}$$

The options are given in millicoulombs (mC). We convert our answer:
Since 1 mC $= 10^{-3}$ C, we have:

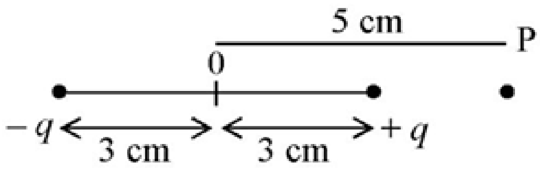
$$q = 2 \text{ mC}$$

Step 4: Final Answer:
The magnitude of the charge on the dipole is 2 mC.

Quick Tip

Always check the units of all given quantities before calculation. In this problem, the dipole length is in cm and must be converted to meters to be consistent with the other SI units (Nm, N/C). Also, be ready to convert the final answer to the unit prefix given in the options (e.g., C to mC).

36. An electric dipole is placed as shown in the figure.



The electric potential (in 10^2 V) at point P due to the dipole is (ϵ_0 = permittivity of free space and $\frac{1}{4\pi\epsilon_0} = K$):

(A) $(\frac{8}{5})$ qK
(B) $(\frac{8}{3})$ qK
(C) $(\frac{3}{8})$ qK
(D) $(\frac{5}{8})$ qK

Correct Answer: (C) $(\frac{3}{8})$ qK

Solution:

Step 1: Understanding the Question:
The problem asks for the electric potential at point P, which lies on the axial line of an electric dipole. We are given the positions of the charges and the point P.

Step 2: Key Formula or Approach:
The electric potential is a scalar quantity. The total potential at a point due to a system of charges is the algebraic sum of the potentials due to individual charges. The potential V at a distance r from a point charge q is given by $V = \frac{Kq}{r}$.

Step 3: Detailed Explanation:
Let's find the distances of point P from the two charges +q and -q.
The center of the dipole is at O. The distance of each charge from the center is 3 cm.
The distance of point P from the center O is 5 cm.
- The distance of point P from the positive charge (+q) is $r_+ = 5 \text{ cm} - 3 \text{ cm} = 2 \text{ cm}$.
- The distance of point P from the negative charge (-q) is $r_- = 5 \text{ cm} + 3 \text{ cm} = 8 \text{ cm}$.

The net potential at point P is the sum of the potentials due to +q and -q.

$$V_P = V_{+q} + V_{-q}$$
$$V_P = \frac{K(+q)}{r_+} + \frac{K(-q)}{r_-} = Kq\left(\frac{1}{r_+} - \frac{1}{r_-}\right)$$

To avoid unit conversion issues, let's keep the distances in cm for the ratio calculation, as the units will cancel out.

$$V_P = Kq\left(\frac{1}{2} - \frac{1}{8}\right) = Kq\left(\frac{4-1}{8}\right) = Kq\left(\frac{3}{8}\right)$$

The question asks for the potential in units of 10^2 V, and the formula uses SI units. Let's verify the units. If distances are in meters, $r_+ = 0.02$ m and $r_- = 0.08$ m.

$$V_P = Kq\left(\frac{1}{0.02} - \frac{1}{0.08}\right) = Kq\,(50 - 12.5) = 37.5Kq = \frac{75}{2}Kq$$

The expression derived using cm is $(\frac{3}{8})\frac{Kq}{\text{cm}}$. To convert this to volts, we must multiply by 100, since 1 cm $= 10^{-2}$ m. So, $V_P = 100 \times (\frac{3}{8})Kq$. The question asks for the value in 10^2 V, which means we need to find the coefficient of 100 V.

$$V_P = \left(\frac{3}{8}qK\right) \times 10^2 \text{ V}$$

So the value is $(\frac{3}{8})qK$.

Step 4: Final Answer:
The electric potential at point P is $(\frac{3}{8})qK$ in units of 10^2 V. This corresponds to option (C).

Quick Tip

When calculating potential (a scalar), you can add the contributions from each charge algebraically. For axial points of a dipole, a shortcut formula is $V = \frac{Kp}{r^2-a^2}$, where p=q(2a). Here, a=3cm, r=5cm, giving $V = \frac{K(q\cdot 6)}{5^2-3^2} = \frac{6Kq}{16} = \frac{3}{8}Kq$. Note that this gives the expression for potential/100, assuming cm are used.

---

37. Two thin lenses are of same focal lengths (f), but one is convex and the other one is concave. When they are placed in contact with each other, the equivalent focal length of the combination will be :
(A) f/2
(B) Infinite
(C) Zero
(D) f/4

Correct Answer: (B) Infinite

Solution:

Step 1: Understanding the Question:
The question asks for the equivalent focal length of a system of two lenses in contact: one convex and one concave, with the same magnitude of focal length.

Step 2: Key Formula or Approach:
For thin lenses in contact, the power of the combination is the sum of the individual powers. The equivalent focal length (f_{eq}) is given by the formula:

$$\frac{1}{f_{eq}} = \frac{1}{f_1} + \frac{1}{f_2}$$

By convention, the focal length of a convex lens is positive, and that of a concave lens is negative.

Step 3: Detailed Explanation:

Let the convex lens be lens 1 and the concave lens be lens 2.

- Focal length of the convex lens, $f_1 = +f$.
- Focal length of the concave lens, $f_2 = -f$.

Now, we apply the formula for the combination:

$$\frac{1}{f_{eq}} = \frac{1}{+f} + \frac{1}{-f}$$
$$\frac{1}{f_{eq}} = \frac{1}{f} - \frac{1}{f} = 0$$

If the reciprocal of the focal length is zero, the focal length itself must be infinite.

$$f_{eq} = \frac{1}{0} = \infty$$

A combination with zero power (P = 1/f) and infinite focal length behaves like a plane glass slab, causing no net convergence or divergence of parallel light rays.

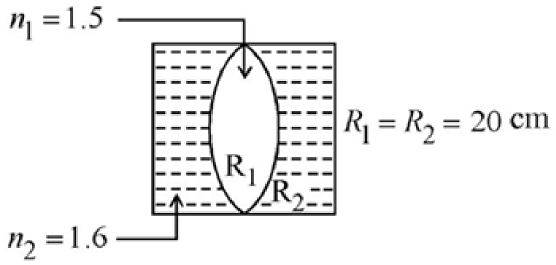
Step 4: Final Answer:

The equivalent focal length of the combination is Infinite. This corresponds to option (B).

> Quick Tip
>
> Remember that power (in Diopters) is the reciprocal of the focal length (in meters). For lenses in contact, powers simply add up. Here, the powers are $+1/f$ and $-1/f$, so the total power is zero, implying infinite focal length.

---

38. In the figure shown here, what is the equivalent focal length of the combination of lenses (Assume that all layers are thin)?



$n_1 = 1.5$, $n_2 = 1.6$, $R_1 = R_2 = 20$ **cm**

(A) -100 cm

(B) - 50 cm
(C) 40 cm
(D) -40 cm

Correct Answer: (A) -100 cm

Solution:

Step 1: Understanding the Question:
The problem asks for the equivalent focal length of a composite lens system. The system consists of three thin lenses in contact: a central biconcave lens made of material n_2 sandwiched between two plano-convex lenses made of material n_1.

Step 2: Key Formula or Approach:
We will use the Lens Maker's formula for each part and then combine their powers.
Lens Maker's formula: $\frac{1}{f} = (n-1)\left(\frac{1}{R_1} - \frac{1}{R_2}\right)$ (assuming the surrounding medium is air, n=1).
For lenses in contact, the equivalent power is the sum of individual powers: $P_{eq} = P_1 + P_2 + P_3$.
The equivalent focal length is $f_{eq} = \frac{1}{P_{eq}}$.
Sign convention: Light is incident from the left. Radii are positive if the center of curvature is to the right of the surface, and negative if it's to the left.

Step 3: Detailed Explanation:
The system can be treated as three lenses in contact:
1. **Lens 1 (Left):** Plano-convex lens of refractive index $n_1 = 1.5$.
Surfaces: Left surface is convex ($R_1 = +20$ cm), right surface is plane ($R_2 = \infty$).
Power $P_1 = (1.5-1)\left(\frac{1}{20} - \frac{1}{\infty}\right) = 0.5 \times \frac{1}{20} = \frac{1}{40}$ D (if distances in m).
2. **Lens 2 (Middle):** Biconcave lens of refractive index $n_2 = 1.6$.
Surfaces: Left surface is concave ($R_1 = -20$ cm), right surface is concave ($R_2 = +20$ cm).
Power $P_2 = (1.6-1)\left(\frac{1}{-20} - \frac{1}{20}\right) = 0.6\left(\frac{-2}{20}\right) = 0.6 \times \frac{-1}{10} = -0.06 = \frac{-6}{100}$ D.
3. **Lens 3 (Right):** Plano-convex lens of refractive index $n_1 = 1.5$.
Surfaces: Left surface is plane ($R_1 = \infty$), right surface is convex ($R_2 = -20$ cm).
Power $P_3 = (1.5-1)\left(\frac{1}{\infty} - \frac{1}{-20}\right) = 0.5 \times \frac{1}{20} = \frac{1}{40}$ D.

Now, calculate the equivalent power (keeping calculations in cm¹ for convenience):

$$\frac{1}{f_{eq}} = P_{eq} = P_1 + P_2 + P_3 = \frac{1}{40} + \left(\frac{-6}{100}\right) + \frac{1}{40}$$
$$\frac{1}{f_{eq}} = \frac{2}{40} - \frac{6}{100} = \frac{1}{20} - \frac{3}{50}$$

Taking a common denominator of 100:

$$\frac{1}{f_{eq}} = \frac{5}{100} - \frac{6}{100} = \frac{-1}{100} \text{ cm}^{-1}$$

Therefore, the equivalent focal length is:

$$f_{eq} = -100 \text{ cm}$$

Step 4: Final Answer:
The equivalent focal length of the combination is -100 cm. This corresponds to option (A).

Quick Tip

When dealing with complex lens combinations, break the system down into simpler, individual lenses in contact. Calculate the power of each lens separately using the Lens Maker's formula with careful attention to sign conventions, then simply add the powers to find the total power of the system.

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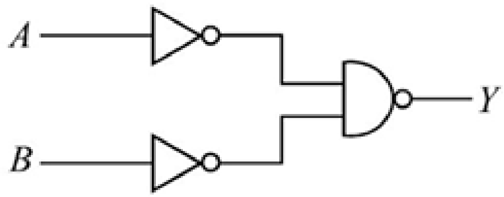
39. For the following logic circuit, the truth table is:



(A) A B Y: 001, 010, 101, 110
(B) A B Y: 000, 010, 100, 111
(C) A B Y: 001, 011, 101, 110
(D) A B Y: 000, 011, 101, 111

Correct Answer: (D) A B Y: 000, 011, 101, 111

Solution:

Step 1: Understanding the Question:
The question asks to determine the truth table for a given digital logic circuit.

Step 2: Key Formula or Approach:
We need to analyze the circuit step-by-step. The circuit consists of a NOT gate and a NAND gate.
- The output of a NOT gate is the inverse of its input: $A' = \text{NOT } A$.
- The output of a NAND gate is the inverse of the AND operation of its inputs: $Y = \text{NOT } (X \cdot Z)$.

Step 3: Detailed Explanation:
Analyzing the given circuit:
1. Input A passes through a NOT gate, so its output is A'.
2. Input B is fed directly into the second gate.
3. The two inputs to the NAND gate are A' and B.
4. The final output Y is therefore $Y = (A' \cdot B)'$.

Using De Morgan's theorem, we can simplify the expression:

$$Y = (A')' + B' = A + B'$$

Let's construct the truth table for this expression, $Y = A + B'$:
- If A=0, B=0: $Y = 0 + 0' = 0 + 1 = 1$.
- If A=0, B=1: $Y = 0 + 1' = 0 + 0 = 0$.
- If A=1, B=0: $Y = 1 + 0' = 1 + 1 = 1$.
- If A=1, B=1: $Y = 1 + 1' = 1 + 0 = 1$.

The correct truth table for the given circuit is: A=0, B=0, Y=1, A=0, B=1, Y=0, A=1, B=0, Y=1, A=1, B=1, Y=1.

Analyzing the Solution:
The answer key indicates option (D) is correct. The truth table for option (D) is: 00→0, 01→1, 10→1, 11→1. This is the truth table for an OR gate ($Y = A + B$).
An OR gate can be implemented using NAND gates if both inputs are inverted before entering the NAND gate: $Y = (A' \cdot B')' = A'' + B'' = A + B$.
This means the circuit diagram likely has a typographical error, and there should have been a NOT gate on the B input line as well. Assuming this correction, the circuit would indeed function as an OR gate, making option (D) the correct answer for the intended circuit.

> **Quick Tip**
>
> If your analysis of a logic circuit leads to a truth table that doesn't match any option, double-check your work. If it's still inconsistent, consider common typographical errors in circuit diagrams, such as a missing gate. In this case, realizing that option (D) is a basic OR gate allows you to hypothesize that the circuit was meant to be a NAND-based OR gate.

---

40. The resistance of platinum wire at 0°C is 2Ω and 6.8Ω at 80°C. The temperature coefficient of resistance of the wire is :
(A) 3×10^{-2} °C^{-1}
(B) 3×10^{-1} °C^{-1}
(C) 3×10^{-4} °C^{-1}
(D) 3×10^{-3} °C^{-1}

Correct Answer: (A) 3×10^{-2} °C^{-1}

Solution:

Step 1: Understanding the Question:
The question asks to find the temperature coefficient of resistance (α) for a platinum wire, given its resistance at two different temperatures.

Step 2: Key Formula or Approach:
The relationship between resistance and temperature is given by the formula:

$$R_T = R_0(1 + \alpha\Delta T)$$

where R_T is the resistance at temperature T, R_0 is the resistance at a reference temperature (here, 0°C), α is the temperature coefficient of resistance, and ΔT is the change in temperature.

Step 3: Detailed Explanation:
We are given:
- Resistance at 0°C, $R_0 = 2\,\Omega$.
- Resistance at 80°C, $R_T = 6.8\,\Omega$.
- The change in temperature, $\Delta T = 80C - 0C = 80C$.

Substitute these values into the formula and solve for α:

$$6.8 = 2(1 + \alpha \times 80)$$

Divide both sides by 2:

$$3.4 = 1 + 80\alpha$$

Subtract 1 from both sides:

$$2.4 = 80\alpha$$

Solve for α:

$$\alpha = \frac{2.4}{80} = \frac{24}{800} = \frac{3}{100}$$
$$\alpha = 0.03\,^{\circ}\mathrm{C}^{-1}$$

In scientific notation, this is:

$$\alpha = 3 \times 10^{-2}\,^{\circ}\mathrm{C}^{-1}$$

Step 4: Final Answer:
The temperature coefficient of resistance of the wire is 3×10^{-2} $^{\circ}\mathrm{C}^{-1}$. This corresponds to option (A).

> **Quick Tip**
>
> The formula $R_T = R_0(1 + \alpha\Delta T)$ is fundamental for problems involving the thermal properties of resistors. Make sure to use the change in temperature (ΔT) correctly.

---

41. A horizontal bridge is built across a river. A student standing on the bridge throws a small ball vertically upwards with a velocity 4 m s¹. The ball strikes the water surface after 4 s. The height of bridge above water surface is (Take g = 10

m s²):
(A) 64 m
(B) 68 m
(C) 56 m
(D) 60 m

Correct Answer: (A) 64 m

Solution:

Step 1: Understanding the Question:
A ball is thrown upwards from a bridge and lands in the water below. We need to find the height of the bridge using the initial velocity, total time of flight, and acceleration due to gravity.

Step 2: Key Formula or Approach:
We can use the second equation of motion for displacement under constant acceleration:

$$s = ut + \frac{1}{2}at^2$$

where s is the net displacement, u is the initial velocity, t is the time, and a is the acceleration. We need to establish a consistent sign convention.

Step 3: Detailed Explanation:
Let's define the point of throw (on the bridge) as the origin (s=0) and take the upward direction as positive.
- Initial velocity, $u = +4$ m/s (since it's thrown upwards).
- Acceleration, $a = -g = -10$ m/s² (since gravity acts downwards).
- Total time, $t = 4$ s.
The final position of the ball is the water surface, which is below the bridge. Let the height of the bridge be h. The net displacement s from the origin to the final position is therefore $-h$.
Now, substitute these values into the equation of motion:

$$s = ut + \frac{1}{2}at^2$$

$$-h = (+4)(4) + \frac{1}{2}(-10)(4)^2$$

$$-h = 16 - 5(16)$$

$$-h = 16 - 80$$

$$-h = -64$$

$$h = 64 \text{ m}$$

Step 4: Final Answer:
The height of the bridge above the water surface is 64 m. This corresponds to option (A).

Quick Tip

In projectile motion problems, choosing a sign convention and sticking to it is crucial. Defining the starting point as the origin and upward as positive is a common and effective strategy. The final displacement will be negative if the object ends up below the starting point.

---

42. A satellite is orbiting just above the surface of the earth with period T. If d is the density of the earth and G is the universal constant of gravitation, the quantity $\frac{3\pi}{Gd}$ represents :

(A) T^3
(B) $\sqrt{T}$
(C) T
(D) T^2

Correct Answer: (D) T^2

Solution:

Step 1: Understanding the Question:
The question asks to identify a physical quantity represented by the expression $\frac{3\pi}{Gd}$, in the context of a satellite orbiting close to the Earth's surface.

Step 2: Key Formula or Approach:
For a satellite in a stable orbit, the gravitational force provides the necessary centripetal force.

$$F_g = F_c$$
$$\frac{GMm}{R^2} = m\omega^2 R$$

where M and R are the mass and radius of the Earth, m is the mass of the satellite, and ω is the angular velocity of the satellite. The angular velocity is related to the time period T by $\omega = \frac{2\pi}{T}$. The mass of the Earth can be expressed in terms of its density d as $M = \text{Volume} \times \text{Density} = \frac{4}{3}\pi R^3 d$.

Step 3: Detailed Explanation:
From the force balance equation:

$$\frac{GM}{R^2} = \omega^2 R \implies GM = \omega^2 R^3$$

Substitute $\omega = \frac{2\pi}{T}$:

$$GM = \left(\frac{2\pi}{T}\right)^2 R^3 = \frac{4\pi^2 R^3}{T^2}$$

Now, substitute the expression for the mass of the Earth, $M = \frac{4}{3}\pi R^3 d$:

$$G\left(\frac{4}{3}\pi R^3 d\right) = \frac{4\pi^2 R^3}{T^2}$$

We can cancel out $4\pi R^3$ from both sides of the equation:

$$G\left(\frac{d}{3}\right) = \frac{\pi}{T^2}$$

Now, we need to rearrange this equation to match the expression given in the question, $\frac{3\pi}{Gd}$. From our derived relation, let's solve for T^2:

$$T^2 = \frac{3\pi}{Gd}$$

This shows that the given quantity represents the square of the orbital period.

Step 4: Final Answer:
The quantity $\frac{3\pi}{Gd}$ represents T^2. This corresponds to option (D).

Quick Tip

This is a classic derivation related to Kepler's Third Law. Whenever you see orbital period (T) and density (d) in a gravitation problem, the key is to relate them by substituting the mass of the central body (M) with its expression in terms of density and volume ($M = d \times V$).

---

43. The radius of inner most orbit of hydrogen atom is 5.3×10^{-11} m. What is the radius of third allowed orbit of hydrogen atom?
(A) 1.59 Å
(B) 4.77 Å
(C) 0.53 Å
(D) 1.06 Å

Correct Answer: (B) 4.77 Å

Solution:

Step 1: Understanding the Question:
The problem provides the radius of the first orbit (ground state) of a hydrogen atom, known as the Bohr radius, and asks for the radius of the third orbit.

Step 2: Key Formula or Approach:
According to the Bohr model for the hydrogen atom, the radius of the n-th allowed orbit is

directly proportional to the square of the principal quantum number (n). The formula is:

$$r_n = r_1 \cdot n^2$$

where r_n is the radius of the n-th orbit and r_1 is the radius of the first orbit (Bohr radius).

Step 3: Detailed Explanation:
We are given:
- Radius of the first orbit, $r_1 = 5.3 \times 10^{-11}$ m.
- We need to find the radius of the third orbit, so $n = 3$.

Using the formula:

$$r_3 = r_1 \cdot (3)^2 = r_1 \cdot 9$$
$$r_3 = (5.3 \times 10^{-11} \text{ m}) \times 9$$
$$r_3 = 47.7 \times 10^{-11} \text{ m}$$

The options are given in Angstroms (Å). We know that 1 Å $= 10^{-10}$ m.
To convert our result to Angstroms, we can write:

$$r_3 = 4.77 \times 10^{-10} \text{ m} = 4.77 \text{ Å}$$

Step 4: Final Answer:
The radius of the third allowed orbit of the hydrogen atom is 4.77 Å. This corresponds to option (B).

Quick Tip

Remember the key dependencies in the Bohr model for hydrogen: Radius $r_n \propto n^2$ and Energy $E_n \propto -1/n^2$. Knowing these proportionalities is often sufficient to solve ratio-based problems quickly.

---

44. A wire carrying a current I along the positive x-axis has length L. It is kept in a magnetic field $\vec{B} = (2\hat{i} + 3\hat{j} - 4\hat{k})$ T. The magnitude of the magnetic force acting on the wire is :

(A) 5 IL
(B) $\sqrt{3}$ IL
(C) 3 IL
(D) $\sqrt{5}$ IL

Correct Answer: (A) 5 IL

Solution:

Step 1: Understanding the Question:
We need to find the magnitude of the magnetic force on a straight current-carrying wire placed in a uniform magnetic field.

Step 2: Key Formula or Approach:
The magnetic force $\vec{F}$ on a wire of length vector $\vec{L}$ carrying current I in a magnetic field $\vec{B}$ is given by the formula:

$$\vec{F} = I(\vec{L} \times \vec{B})$$

The magnitude of this force is $|\vec{F}|$.

Step 3: Detailed Explanation:
First, we define the length vector $\vec{L}$. The wire has length L and is along the positive x-axis. So, the vector is:

$$\vec{L} = L\hat{\imath}$$

The magnetic field is given as:

$$\vec{B} = (2\hat{\imath} + 3\hat{\jmath} - 4\hat{k}) \text{ T}$$

Now, we calculate the cross product $\vec{L} \times \vec{B}$:

$$\vec{L} \times \vec{B} = (L\hat{\imath}) \times (2\hat{\imath} + 3\hat{\jmath} - 4\hat{k})$$
$$\vec{L} \times \vec{B} = L(2)(\hat{\imath} \times \hat{\imath}) + L(3)(\hat{\imath} \times \hat{\jmath}) - L(4)(\hat{\imath} \times \hat{k})$$

Using the properties of cross products of unit vectors ($\hat{\imath} \times \hat{\imath} = 0$, $\hat{\imath} \times \hat{\jmath} = \hat{k}$, $\hat{\imath} \times \hat{k} = -\hat{\jmath}$):

$$\vec{L} \times \vec{B} = L(0) + 3L(\hat{k}) - 4L(-\hat{\jmath})$$
$$\vec{L} \times \vec{B} = 3L\hat{k} + 4L\hat{\jmath}$$

The force vector is:

$$\vec{F} = I(4L\hat{\jmath} + 3L\hat{k}) = IL(4\hat{\jmath} + 3\hat{k})$$

The magnitude of the force is:

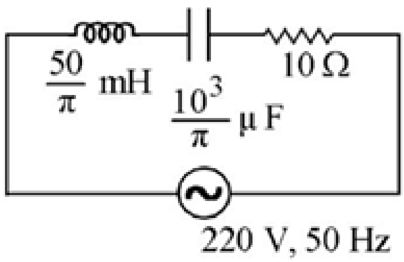
$$|\vec{F}| = |IL(4\hat{\jmath} + 3\hat{k})| = IL\sqrt{4^2 + 3^2}$$
$$|\vec{F}| = IL\sqrt{16 + 9} = IL\sqrt{25} = 5IL$$

Step 4: Final Answer:
The magnitude of the magnetic force acting on the wire is 5 IL.

Quick Tip

Remember the formula for magnetic force is $\vec{F} = I(\vec{L} \times \vec{B})$. Pay close attention to the direction of the length vector $\vec{L}$, which is in the direction of the current. The cross product properties of unit vectors are essential for such problems.

45. The net impedance of circuit (as shown in figure) will be :



(A) $5\sqrt{5}\Omega$
(B) 25Ω
(C) $10\sqrt{2}\Omega$
(D) 15Ω

Correct Answer: (A) $5\sqrt{5}\Omega$

Solution:

Step 1: Understanding the Question:
The question asks for the net impedance (Z) of a series LCR circuit. We are given the values of the inductor (L), capacitor (C), resistor (R), and the frequency (f) of the AC source.

Step 2: Key Formula or Approach:
The impedance Z of a series LCR circuit is given by:

$$Z = \sqrt{R^2 + (X_L - X_C)^2}$$

where X_L is the inductive reactance and X_C is the capacitive reactance.
Inductive reactance: $X_L = 2\pi f L$
Capacitive reactance: $X_C = \frac{1}{2\pi f C}$

Step 3: Detailed Explanation:
First, list the given values in SI units:
Resistance, $R = 10\,\Omega$
Inductance, $L = \frac{50}{\pi}$ mH $= \frac{50}{\pi} \times 10^{-3}$ H
Capacitance, $C = \frac{10^3}{\pi}\,\mu\text{F} = \frac{10^3}{\pi} \times 10^{-6}$ F $= \frac{10^{-3}}{\pi}$ F
Frequency, $f = 50$ Hz

Next, calculate the inductive reactance X_L:

$$X_L = 2\pi f L = 2\pi \times 50 \times \left(\frac{50}{\pi} \times 10^{-3}\right) = 100\pi \times \frac{50}{\pi} \times 10^{-3} = 5000 \times 10^{-3} = 5\,\Omega$$

Next, calculate the capacitive reactance X_C:

$$X_C = \frac{1}{2\pi f C} = \frac{1}{2\pi \times 50 \times \left(\frac{10^{-3}}{\pi}\right)} = \frac{1}{100\pi \times \frac{10^{-3}}{\pi}} = \frac{1}{100 \times 10^{-3}} = \frac{1}{10^{-1}} = 10\,\Omega$$

Now, calculate the impedance Z:

$$Z = \sqrt{R^2 + (X_L - X_C)^2} = \sqrt{10^2 + (5 - 10)^2}$$

$$Z = \sqrt{100 + (-5)^2} = \sqrt{100 + 25} = \sqrt{125}$$

$$Z = \sqrt{25 \times 5} = 5\sqrt{5}\,\Omega$$

Step 4: Final Answer:
The net impedance of the circuit is $5\sqrt{5}\,\Omega$.

> **Quick Tip**
>
> In LCR circuit problems, always calculate inductive reactance (X_L) and capacitive reactance (X_C) first. Then substitute these values into the impedance formula. Be careful with the units (mH, μF) and convert them to base SI units (H, F) before calculation.

---

46. 10 resistors, each of resistance R are connected in series to a battery of emf E and negligible internal resistance. Then those are connected in parallel to the same battery, the current is increased n times. The value of n is:

(A) 1
(B) 1000
(C) 10
(D) 100

Correct Answer: (D) 100

Solution:

Step 1: Understanding the Question:
We have two scenarios with 10 identical resistors and a battery. First, they are connected in series, and then in parallel. We are told the current in the parallel connection is 'n' times the current in the series connection, and we need to find 'n'.

Step 2: Key Formula or Approach:
We need the formulas for equivalent resistance for series and parallel combinations, and Ohm's Law ($I = V/R_{eq}$).
For 'N' identical resistors of resistance R:
- Series equivalent resistance: $R_S = N \times R$

- Parallel equivalent resistance: $R_P = R/N$
The current is given by $I = E/R_{eq}$, where E is the emf of the battery.

Step 3: Detailed Explanation:
Number of resistors, N = 10.
Case 1: Series Connection
The equivalent resistance is:

$$R_S = 10 \times R = 10R$$

The current flowing from the battery is:

$$I_S = \frac{E}{R_S} = \frac{E}{10R}$$

Case 2: Parallel Connection
The equivalent resistance is:

$$R_P = \frac{R}{10}$$

The current flowing from the battery is:

$$I_P = \frac{E}{R_P} = \frac{E}{(R/10)} = \frac{10E}{R}$$

Relating the Currents
We are given that the current is increased n times, which means $I_P = n \times I_S$.

$$\frac{10E}{R} = n \times \left(\frac{E}{10R}\right)$$

We can cancel E and R from both sides:

$$10 = n \times \frac{1}{10}$$

Solving for n:

$$n = 10 \times 10 = 100$$

Step 4: Final Answer:
The value of n is 100.

Quick Tip

For N identical resistors, the ratio of series to parallel equivalent resistance is $R_S/R_P = (NR)/(R/N) = N^2$. Since current is inversely proportional to resistance ($I \propto 1/R$), the ratio of currents will be the inverse: $I_P/I_S = R_S/R_P = N^2$. Here N=10, so n = $10^2 = 100$.

---

47. Calculate the maximum acceleration of a moving car so that a body lying on the floor of the car remains stationary. The coefficient of static friction between the body and the floor is 0.15 (g = 10 m s^{-2}).

(A) 1.5 ms^{-2}
(B) 50 ms^{-2}
(C) 1.2 ms^{-2}
(D) 150 m s^{-2}

Correct Answer: (A) 1.5 ms^{-2}

Solution:

Step 1: Understanding the Question:
A body is on the floor of an accelerating car. For the body to not slip (remain stationary relative to the car), the force causing it to accelerate must be provided by static friction. We need to find the maximum possible acceleration.

Step 2: Key Formula or Approach:
The force required to accelerate the body of mass 'm' with an acceleration 'a' is given by Newton's second law: $F = ma$.
This force is provided by the static friction, f_s.
The maximum value of static friction is given by $f_{s,\max} = \mu_s N$, where μ_s is the coefficient of static friction and N is the normal force. On a horizontal floor, $N = mg$.
For the body to remain stationary, the required force must be less than or equal to the maximum static friction: $ma \leq f_{s,\max}$.

Step 3: Detailed Explanation:
Let 'm' be the mass of the body.
The force needed to accelerate the body with the car is $F = ma$.
This force is provided by static friction, f_s. So, $f_s = ma$.
The maximum available static friction force is:

$$f_{s,\max} = \mu_s N = \mu_s mg$$

For the body to not slip, the required force cannot exceed the maximum available frictional force:

$$ma \leq \mu_s mg$$

Dividing by 'm' on both sides, we get the condition for the acceleration:

$$a \leq \mu_s g$$

The maximum possible acceleration $a_{\max}$ is therefore:

$$a_{\max} = \mu_s g$$

Substitute the given values:
$\mu_s = 0.15$
$g = 10$ m/s^2

$$a_{\text{max}} = 0.15 \times 10 = 1.5 \text{ m/s}^2$$

Step 4: Final Answer:
The maximum acceleration of the car is 1.5 ms^{-2}.

Quick Tip

This is a classic problem of a non-inertial frame. The static friction provides the pseudo-force's counteraction. A simple way to remember is that the maximum acceleration an object can have on a surface due to friction is $a_{\text{max}} = \mu g$. This applies to both static and kinetic friction depending on the problem.

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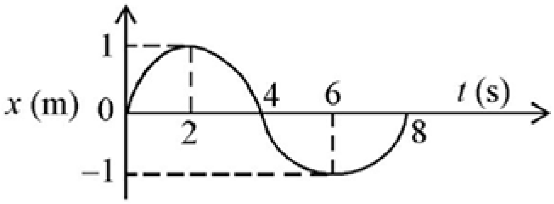
48. The x-t graph of a particle performing simple harmonic motion is shown in the figure. The acceleration of the particle at t=2s is:



(A) $\frac{\pi^2}{16}$ ms^{-2}
(B) $\frac{\pi^2}{8}$ ms^{-2}
(C) $-\frac{\pi^2}{8}$ ms^{-2}
(D) $-\frac{\pi^2}{16}$ ms^{-2}

Correct Answer: (D) $-\frac{\pi^2}{16}$ ms^{-2}

Solution:

Step 1: Understanding the Question:
We are given the position-time (x-t) graph for a particle in Simple Harmonic Motion (SHM) and asked to find its acceleration at a specific time, t = 2 s.

Step 2: Key Formula or Approach:
In SHM, the acceleration a is related to the position x by the formula:

$$a(t) = -\omega^2 x(t)$$

where ω is the angular frequency. The angular frequency is related to the time period T by:

$$\omega = \frac{2\pi}{T}$$

We can determine the amplitude A, time period T, and the position x at t=2s from the given graph.

Step 3: Detailed Explanation:
First, let's analyze the graph to find the parameters of the motion.
- **Amplitude (A):** The maximum displacement from the mean position. From the graph, the maximum value of x is 1 m. So, $A = 1$ m.
- **Time Period (T):** The time taken for one complete oscillation. The graph shows one full cycle from t=0 to t=8 s (it goes from 0 to 1, back to 0, to -1, and back to 0). So, $T = 8$ s.

Next, calculate the angular frequency ω:

$$\omega = \frac{2\pi}{T} = \frac{2\pi}{8} = \frac{\pi}{4} \text{ rad/s}$$

Now, we need the position of the particle at t = 2 s. Looking at the graph, at t = 2 s, the particle is at its maximum positive displacement.

$$x(t = 2\text{s}) = +1 \text{ m}$$

Finally, calculate the acceleration at t = 2 s using the formula $a = -\omega^2 x$:

$$a(t = 2\text{s}) = -\left(\frac{\pi}{4}\right)^2 \times x(t = 2\text{s})$$

$$a(t = 2\text{s}) = -\frac{\pi^2}{16} \times (1) = -\frac{\pi^2}{16} \text{ m/s}^2$$

Step 4: Final Answer:
The acceleration of the particle at t = 2 s is $-\frac{\pi^2}{16}$ ms^{-2}.

Quick Tip

In SHM, acceleration is maximum in magnitude at the extreme positions (where displacement is maximum) and is zero at the mean position. The direction of acceleration is always opposite to the direction of displacement. At the positive extreme (x=+A), acceleration is maximum negative ($a = -\omega^2 A$).

---

49. A bullet from a gun is fired on a rectangular wooden block with velocity u. When bullet travels 24 cm through the block along its length horizontally, velocity of bullet becomes u/3. Then it further penetrates into the block in the same direction before coming to rest exactly at the other end of the block. The total length of the block is :

(A) 28 cm
(B) 30 cm
(C) 27 cm
(D) 24 cm

Correct Answer: (C) 27 cm

Solution:

Step 1: Understanding the Question:
A bullet penetrates a wooden block, and its velocity decreases. We are given its initial velocity, its velocity after traveling 24 cm, and the fact that it stops at the end of the block. We need to find the total length of the block, assuming constant deceleration.

Step 2: Key Formula or Approach:
We will use the third equation of motion, assuming constant acceleration (in this case, deceleration 'a'):

$$v^2 = u^2 + 2as$$

where v is the final velocity, u is the initial velocity, a is the acceleration, and s is the distance traveled. We apply this equation to two parts of the bullet's journey.

Step 3: Detailed Explanation:
Let the constant deceleration be a. The motion can be divided into two parts.

Part 1: The bullet travels 24 cm.
Initial velocity, $u_1 = u$
Final velocity, $v_1 = u/3$
Distance, $s_1 = 24$ cm
Using the equation of motion:

$$(u/3)^2 = u^2 + 2as_1$$
$$\frac{u^2}{9} = u^2 + 2a(24)$$
$$\frac{u^2}{9} - u^2 = 48a$$
$$-\frac{8u^2}{9} = 48a$$
$$a = -\frac{8u^2}{9 \times 48} = -\frac{u^2}{54}$$

Part 2: The bullet travels the remaining distance s_2 and comes to rest.
Initial velocity, $u_2 = u/3$
Final velocity, $v_2 = 0$
Distance, s_2

Using the equation of motion again with the same deceleration a:

$$v_2^2 = u_2^2 + 2as_2$$
$$0^2 = (u/3)^2 + 2as_2$$
$$0 = \frac{u^2}{9} + 2as_2$$
$$s_2 = -\frac{u^2/9}{2a} = -\frac{u^2}{18a}$$

Now, substitute the value of a we found from Part 1:

$$s_2 = -\frac{u^2}{18(-\frac{u^2}{54})} = \frac{u^2 \times 54}{18u^2} = \frac{54}{18} = 3 \text{ cm}$$

Total Length:
The total length of the block is the sum of the distances from both parts:

$$L = s_1 + s_2 = 24 \text{ cm} + 3 \text{ cm} = 27 \text{ cm}$$

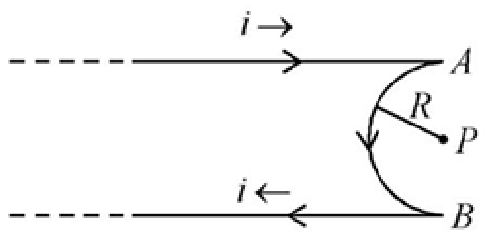
Step 4: Final Answer:
The total length of the block is 27 cm.

Quick Tip

Alternatively, use the work-energy theorem. The work done by the resistive force is equal to the change in kinetic energy. Let the resistive force be F. In the first part, $F\times 24 = \frac{1}{2}m(u^2-(u/3)^2) = \frac{1}{2}m\frac{8u^2}{9}$. In the second part, $F\times s_2 = \frac{1}{2}m((u/3)^2-0) = \frac{1}{2}m\frac{u^2}{9}$. Dividing the two equations gives $\frac{24}{s_2} = 8$, so $s_2 = 3$ cm.

---

50. A very long conducting wire is bent in a semi-circular shape from A to B as shown in figure. The magnetic field at point P for steady current configuration is given by :



(A) $\frac{\mu_0 i}{4R}\left[1 - \frac{2}{\pi}\right]$ pointed away from page
(B) $\frac{\mu_0 i}{4R}\left[1 + \frac{2}{\pi}\right]$ pointed into the page
(C) $\frac{\mu_0 i}{4R}$ pointed into the page
(D) $\frac{\mu_0 i}{4R}$ pointed away from the page

Correct Answer: (A) $\frac{\mu_0 i}{4R}\left[1-\frac{2}{\pi}\right]$ pointed away from page

Solution:

Step 1: Understanding the Question:
We need to find the net magnetic field at point P, which is the center of the semi-circular part of a long wire carrying a current i. The total field is the vector sum of the fields from the two straight sections and the semi-circular section.

Step 2: Key Formula or Approach:
- Magnetic field due to a semi-infinite straight wire at a perpendicular distance R from its end: $B_{\text{semi-straight}} = \frac{\mu_0 i}{4\pi R}$. - Magnetic field at the center of a semi-circular arc of radius R: $B_{\text{arc}} = \frac{\mu_0 i}{4R}$.
- The direction of the magnetic field is determined by the Right-Hand Rule.

Step 3: Detailed Explanation:
Let's analyze the contribution from each of the three parts of the wire.
1. The left straight semi-infinite section:
The current flows to the right. Using the Right-Hand Rule (pointing thumb in the direction of current), the magnetic field at point P (which is above the wire) points **out of the page**. The magnitude is:

$$B_1 = \frac{\mu_0 i}{4\pi R}$$

2. The right straight semi-infinite section:
The current flows to the right. Similarly, the magnetic field at point P also points **out of the page**. The magnitude is:

$$B_3 = \frac{\mu_0 i}{4\pi R}$$

The total field from the two straight sections is:

$$B_{\text{straight}} = B_1 + B_3 = 2 \times \frac{\mu_0 i}{4\pi R} = \frac{\mu_0 i}{2\pi R} \text{ (out of page)}$$

3. The semi-circular arc (A to B):
The current flows clockwise along the arc. Using the Right-Hand Curl Rule (curling fingers in the direction of current), the thumb points **into the page**. The magnitude is:

$$B_{\text{arc}} = \frac{\mu_0 i}{4R} \text{ (into page)}$$

Net Magnetic Field:

The diagram and the current actually flows from right to left ($\leftarrow i$).
- Straight parts: Current left, field at P is **into the page**.

- Arc: Current is now counter-clockwise, field at P is **out of the page**.

The net field would be $B_{\text{net}} = B_{\text{arc}} - B_{\text{straight}} = \frac{\mu_0 i}{4R} - \frac{\mu_0 i}{2\pi R} = \frac{\mu_0 i}{4R}\left(1 - \frac{2}{\pi}\right)$.

Since $1 > 2/\pi$, the result is positive, meaning the direction is that of B_{arc}, which is **out of the page** (away from page).

This matches option (A) perfectly.

Step 4: Final Answer:
Assuming the current direction is opposite to that shown in the diagram (i.e., flowing from right to left), the net magnetic field at P is $\frac{\mu_0 i}{4R}\left[1 - \frac{2}{\pi}\right]$ pointed away from the page.

> **Quick Tip**
>
> When dealing with composite wire shapes, calculate the magnetic field from each segment separately and then perform a vector sum. Always be very careful with the Right-Hand Rule to determine the direction of the field from each segment. If your calculated answer's magnitude matches an option but the direction doesn't, consider the possibility of a mislabeled diagram.

Chemistry

51. Amongst the given options which of the following molecules/ion acts as a Lewis acid?

(A) BF_3
(B) OH^-
(C) NH_3
(D) H_2O

Correct Answer: (A) BF_3

Solution:

Step 1: Understanding the Question:
The question asks to identify the Lewis acid from a given list of species.

Step 2: Key Formula or Approach:
Let's define Lewis acids and bases:

- A **Lewis acid** is a chemical species that is an electron-pair acceptor. Typically, these are species with an incomplete octet of electrons on the central atom or with vacant d-orbitals.
- A **Lewis base** is a chemical species that is an electron-pair donor. Typically, these are species with lone pairs of electrons.

Step 3: Detailed Explanation:
Let's analyze each option:

- **(A) BF_3 (Boron trifluoride):** The central atom is Boron (B). Boron is in Group 13 and has 3 valence electrons. In BF_3, it forms three single bonds with three fluorine atoms. The total number of electrons around the Boron atom is $3 \times 2 = 6$. Since Boron has an incomplete octet, it can accept a pair of electrons to complete its octet. Therefore, BF_3 acts as a **Lewis acid**.
- **(B) OH^- (Hydroxide ion):** The oxygen atom has two lone pairs of electrons and a negative charge. It can readily donate an electron pair to form a coordinate bond. Therefore, OH^- acts as a **Lewis base**.
- **(C) NH_3 (Ammonia):** The central nitrogen atom has one lone pair of electrons. It can donate this electron pair. Therefore, NH_3 acts as a **Lewis base**.
- **(D) H_2O (Water):** The central oxygen atom has two lone pairs of electrons. It can donate one of these electron pairs. Therefore, H_2O acts as a **Lewis base**. It can also act as a Brønsted-Lowry acid, but its primary Lewis character is basic.

Step 4: Final Answer:
Among the given options, only BF_3 is an electron-pair acceptor and thus acts as a Lewis acid.

Quick Tip

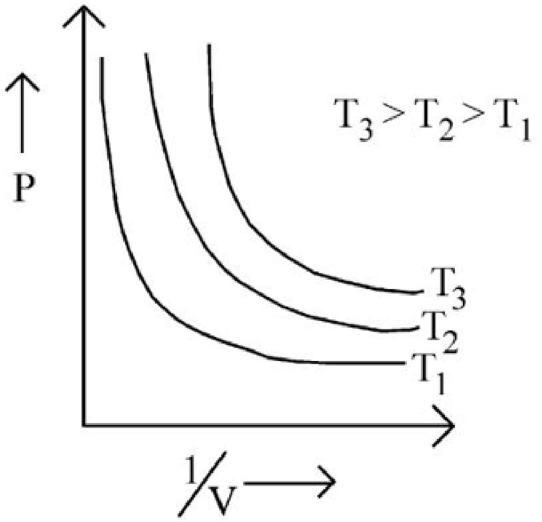
Look for the defining feature:

- Lewis bases almost always have lone pairs of electrons (e.g., NH_3, H_2O, OH^-, Cl^-).
- Lewis acids often have an incomplete octet (e.g., BF_3, $AlCl_3$), are cations (e.g., H^+, Ag^+), or have central atoms with vacant d-orbitals that can accept electrons (e.g., SiF_4, $SnCl_4$).

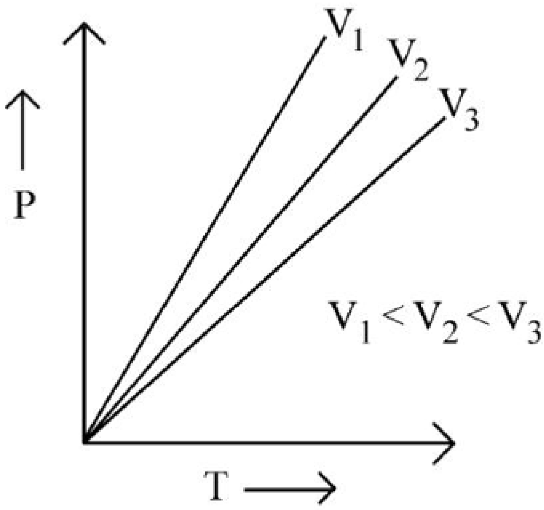
---

52. Which amongst the following options is correct graphical representation of Boyle's Law?

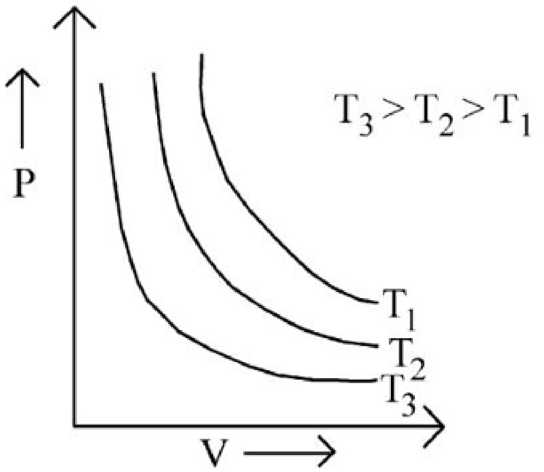
(1)


P
$T_3 > T_2 > T_1$
T_3
T_2
T_1
$1/V$

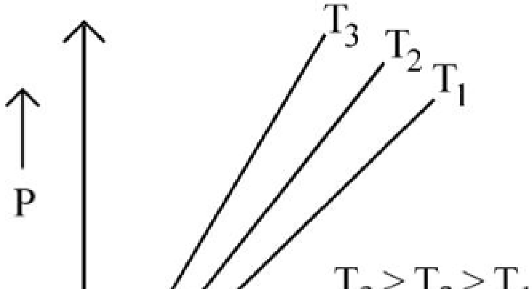
(2)


P
V_1
V_2
V_3
$V_1 < V_2 < V_3$
T

(3)


P
$T_3 > T_2 > T_1$
T_1
T_2
T_3
V

(4)


P
T_3
T_2
T_1

(A) Graph (1)
(B) Graph (2)
(C) Graph (3)
(D) Graph (4)

Correct Answer: (C) Graph (3)

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct graphical representation of Boyle's Law from the given options. Boyle's law describes the relationship between pressure (P) and volume (V) of a gas at constant temperature.

Step 2: Key Formula or Approach:
Boyle's Law states that for a fixed amount of gas at constant temperature, the pressure is inversely proportional to the volume.

$$P \propto \frac{1}{V} \quad \text{or} \quad PV = k \quad \text{(where k is a constant)}$$

From the ideal gas law, $PV = nRT$, we can see that the constant $k = nRT$. This means that for a fixed amount of gas (n), the value of the constant k is directly proportional to the absolute temperature (T).

Step 3: Detailed Explanation:
Let's analyze the graphs provided:

- The graphs plot Pressure (P) on the y-axis and either Volume (V) or 1/V on the x-axis. The different curves on each graph represent different constant temperatures (isotherms).
- **Graph of P vs V:** According to Boyle's law, $P = k/V$. This represents a hyperbolic relationship. The graphs in options (1) and (3) show this hyperbolic shape.
- **Graph of P vs 1/V:** According to Boyle's law, $P = k \times (1/V)$. This is of the form $y = mx$, which is a straight line passing through the origin with a slope of $m = k$. The graphs in options (2) and (4) plot P vs T (incorrect for Boyle's law) or P vs 1/V. Graph (4) shows P vs 1/V as a straight line through the origin, which is correct.

Now let's consider the effect of temperature.

- **For P vs V graphs (1 and 3):** The constant $k = nRT$. As temperature T increases, the value of PV increases. This means that for a given volume V, a higher temperature T will result in a higher pressure P. Therefore, the isotherm for a higher temperature should lie above the isotherm for a lower temperature.
  - In Graph (1), for a fixed V, the pressure on the T_1 curve is the highest, implying T_1 ¿ T_2 ¿ T_3. This contradicts the label T_3 ¿ T_2 ¿ T_1. So, (1) is incorrect.

– In Graph (3), for a fixed V, the pressure on the T_3 curve is the highest, followed by T_2, and then T_1. This corresponds to the relationship T_3 ¿ T_2 ¿ T_1, which is given in the label. So, (3) is the **correct** representation.

- **For P vs 1/V graph (4):** The slope of the line is $k = nRT$. A higher temperature T should result in a steeper slope. In Graph (4), the line for T_3 has the shallowest slope, while T_1 has the steepest. This would mean T_1 ¿ T_2 ¿ T_3. This contradicts the label T_3 ¿ T_2 ¿ T_1. So, (4) is incorrect.
- **Graph (2)** plots P vs T at constant volume (Gay-Lussac's Law), not Boyle's Law.

Step 4: Final Answer:
Graph (3) correctly shows the hyperbolic relationship between P and V, and also correctly depicts the relative positions of the isotherms for different temperatures (higher temperature isotherm is further from the origin).

Quick Tip

To quickly check the temperature dependence on a P-V graph, remember the ideal gas law $PV = nRT$. For a higher T, the product PV must be larger. This means the curve (isotherm) for a higher T must be "further out" from the origin than the curve for a lower T.

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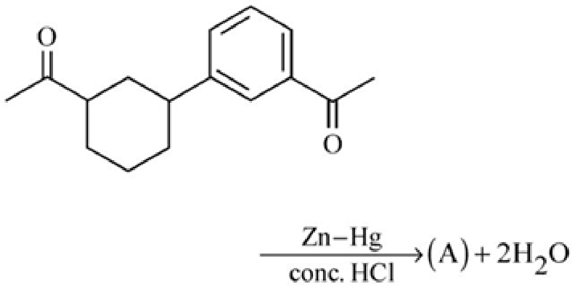
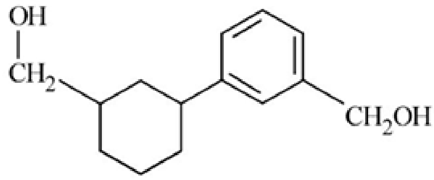
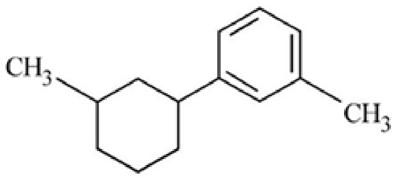
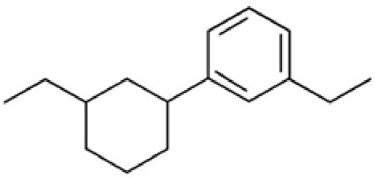
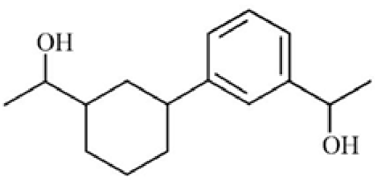
53. Identify product (A) in the following reaction:



(1) 

(2) 

(3) 

(4) 

Correct Answer: (C) Structure with two ethyl groups.

Solution:

Step 1: Understanding the Question:
The question shows a reaction involving a diketone and asks to identify the product (A). We need to recognize the reagents and the transformation they cause.

Step 2: Key Formula or Approach:
The reagents used are Zn-Hg (zinc amalgam) and concentrated HCl. This set of reagents is used for the **Clemmensen reduction**.
The Clemmensen reduction is a chemical reaction used to reduce aldehydes or ketones to the corresponding alkanes. It completely removes the oxygen from the carbonyl group (C=O) and replaces it with two hydrogen atoms (CH_2).

$$\text{R-CO-R'} \xrightarrow{\text{Zn-Hg, conc. HCl}} \text{R-CH}_2\text{-R'}$$

Step 3: Detailed Explanation:
The starting material has two ketone (carbonyl) groups: one on the cyclohexyl ring and one on the benzene ring. Both of these ketone groups will be reduced by the Clemmensen reagents.

- The acetyl group (-$COCH_3$) on the cyclohexyl ring will be reduced to an ethyl group (-CH_2CH_3).
- The acetyl group (-$COCH_3$) on the benzene ring will also be reduced to an ethyl group (-CH_2CH_3).

The overall transformation is:

$$\text{Structure with two -COCH}_3\text{ groups} \xrightarrow{\text{Zn-Hg, conc. HCl}} \text{Structure with two -CH}_2\text{CH}_3\text{ groups}$$

The product (A) will be the original carbon skeleton but with two ethyl groups instead of the two acetyl groups. Let's analyze the options:

- (1) Shows reduction to secondary alcohols followed by dehydration and rearrangement. Incorrect.
- (2) and (3) both show the correct product, which is the starting molecule with both C=O groups reduced to CH_2. So, the acetyl groups become ethyl groups. Let's assume (3) is the intended correct representation.
- (4) Shows reduction of the ketones to secondary alcohols. This would be the product of reduction with a milder reducing agent like $NaBH_4$, not the Clemmensen reduction.

Step 4: Final Answer:
The Clemmensen reduction reduces both ketone groups to methylene groups, converting the two acetyl groups into two ethyl groups. The correct product is shown in option (3).

Quick Tip

Remember the two main reactions for reducing carbonyls to alkanes:

- **Clemmensen Reduction:** Zn-Hg, conc. HCl. Works under acidic conditions. Not suitable for acid-sensitive substrates.
- **Wolff-Kishner Reduction:** NH_2NH_2, KOH, heat. Works under basic conditions. Not suitable for base-sensitive substrates.

Both achieve the same C=O $\rightarrow$ CH_2 transformation.

---

54. Which of the following statements are NOT correct?
A. Hydrogen is used to reduce heavy metal oxides to metals.
B. Heavy water is used to study reaction mechanism.
C. Hydrogen is used to make saturated fats from oils.
D. The H-H bond dissociation enthalpy is lowest as compared to a single bond between two atoms of any element.
E. Hydrogen reduces oxides of metals that are more active than iron.

Choose the most appropriate answer from the options given below :

(A) D, E only
(B) A, B, C only
(C) B, C, D, E only
(D) B, D only

Correct Answer: (A) D, E only

Solution:

Step 1: Understanding the Question:
We need to identify the incorrect statements about hydrogen and its compounds from the given list of five statements.

Step 3: Detailed Explanation:
Let's evaluate each statement:

- **A. Hydrogen is used to reduce heavy metal oxides to metals.**
  This is **correct**. Hydrogen is a good reducing agent and is used in metallurgy to reduce oxides of less reactive metals (like Cu, Pb, Sn) to their respective metals. E.g., $CuO + H_2 \rightarrow Cu + H_2O$.

- **B. Heavy water is used to study reaction mechanism.**
  This is **correct**. Heavy water (D_2O) is used as a tracer to study the mechanism of chemical reactions. By replacing hydrogen with deuterium, chemists can track the path of the hydrogen atoms through the reaction.

- **C. Hydrogen is used to make saturated fats from oils.**
  This is **correct**. This process is called hydrogenation of oils. Unsaturated fats (containing C=C double bonds) in vegetable oils are reacted with hydrogen in the presence of a catalyst (like Ni, Pt, or Pd) to produce saturated fats (like vanaspati ghee).

- **D. The H-H bond dissociation enthalpy is lowest as compared to a single bond between two atoms of any element.**
  This is **false**. The H-H bond dissociation enthalpy is very high (435.88 kJ/mol). It is one of the strongest single bonds known, not the lowest. For example, the C-C bond ($\sim$348 kJ/mol) and Cl-Cl bond ($\sim$242 kJ/mol) are weaker.

- **E. Hydrogen reduces oxides of metals that are more active than iron.**
  This is **false**. Hydrogen can only reduce the oxides of metals that are less reactive than it. In the reactivity series, metals like Na, K, Ca, Mg, Al, and Fe are more reactive than hydrogen. Their oxides are very stable and cannot be reduced by hydrogen. Hydrogen can reduce oxides of metals like copper, lead, and tin, which are less reactive than iron.

The statements that are NOT correct are D and E.

Step 4: Final Answer:
The incorrect statements are D and E. This corresponds to option (A).

Quick Tip

To decide if hydrogen can reduce a metal oxide, use the electrochemical/reactivity series. Hydrogen can generally reduce the oxides of metals placed below it in the series, but not those placed significantly above it.

---

55. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R :
Assertion A : In equation $\Delta_r G = -nFE_{\textbf{cell}}$, value of $\Delta_r G$ depends on n.
Reason R: $E_{\textbf{cell}}$ is an intensive property and $\Delta_r G$ is an extensive property.
In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true and R is NOT the correct explanation of A.

Correct Answer: (D) Both A and R are true and R is NOT the correct explanation of A.

Solution:

Step 1: Understanding the Question:
This is an Assertion-Reason question concerning the relationship between Gibbs free energy ($\Delta_r G$) and cell potential (E_{cell}) in electrochemistry. We need to evaluate the truth of both statements and their relationship.

Step 3: Detailed Explanation:
Analysis of Assertion A:
The assertion states that in the equation $\Delta_r G = -nFE_{\text{cell}}$, the value of $\Delta_r G$ depends on n. Here, n represents the number of moles of electrons transferred in the balanced redox reaction for which $\Delta_r G$ is being calculated. The equation clearly shows that $\Delta_r G$ is directly proportional to n. If you double the stoichiometric coefficients in the balanced reaction, the value of n doubles, and consequently, the value of $\Delta_r G$ also doubles. Therefore, Assertion A is **true**.

Analysis of Reason R:
The reason states that E_{cell} is an intensive property and $\Delta_r G$ is an extensive property.

- An **intensive property** is a property of matter that does not depend on the amount of substance present. Examples include temperature, pressure, and concentration. Cell potential (E_{cell}) is an intensive property because it is a potential difference and does not change if you increase the size of the electrochemical cell or the amount of reactants (as long as concentrations remain the same).
- An **extensive property** is a property that depends on the amount of substance. Examples include mass, volume, and energy. Gibbs free energy ($\Delta_r G$) is an extensive property

because it represents the total energy change for the reaction as written, which is proportional to the amount of substance reacting (represented by n).

Both parts of this statement are correct. Reason R is **true**.

Analysis of the relationship:
Now, we must determine if R is the correct explanation for A.
Assertion A is true: ΔG is directly proportional to n, as seen in the equation. Reason R is true: ΔG is an extensive property (depends on amount) and E_{cell} is an intensive property (does not depend on amount). However, R is not the direct explanation for A. A is a mathematical fact derived from the equation itself. R provides a classification of the terms involved. While related, one isn't the direct cause of the other in a simple sense. The dependence on n makes ΔG extensive, so A is a specific instance of the general property mentioned in R. It's more like A is an example of R, not that R is the reason for A. Hence, R is not the correct explanation of A.

Step 4: Final Answer:
Both Assertion A and Reason R are true statements. However, the reason (a statement about the nature of the properties) is not considered the direct explanation for the assertion (a mathematical dependence evident from the formula itself).

Quick Tip

Remember the difference between intensive and extensive properties. Intensive properties (like potential, density, temperature) are independent of system size. Extensive properties (like energy, mass, volume) are proportional to system size. In $\Delta G = -nFE_{cell}$, the equation correctly balances, as an extensive quantity (ΔG) is equated to the product of an intensive quantity (E_{cell}) and an extensive quantity (nF).

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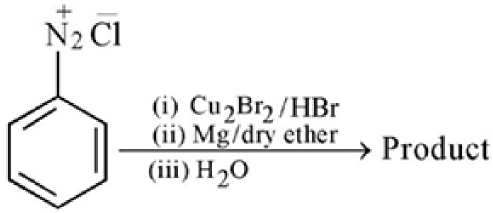
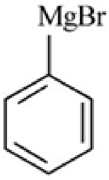
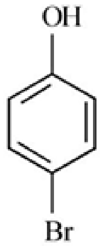
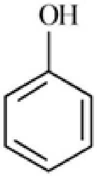
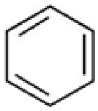
56. Identify the product in the following reaction:



(1) 

(2) 

(3) 

(4) 

Correct Answer: (D) Benzene

Solution:

Step 1: Understanding the Question:
This is a multi-step organic synthesis problem starting from benzenediazonium chloride. We need to follow the sequence of reactions to identify the final product.

Step 2: Key Formula or Approach:
We need to identify the purpose of each reagent in the sequence.

- **Step (i) Cu_2Br_2/HBr:** This is the Sandmeyer reaction, which replaces the diazonium group (-$N_2^+Cl^-$) with a bromine atom.
- **Step (ii) Mg/dry ether:** This reagent is used to form a Grignard reagent from an aryl or alkyl halide.
- **Step (iii) H_2O:** This is the workup step. Adding water to a Grignard reagent protonates it, replacing the -MgBr group with a hydrogen atom.

Step 3: Detailed Explanation:
Let's trace the reaction step-by-step:

- **Starting Material:** Benzenediazonium chloride.
- **Step (i):** Reaction with Cu_2Br_2/HBr.

$$\mathrm{C_6H_5N_2^+Cl^-} \xrightarrow{\mathrm{Cu_2Br_2/HBr}} \mathrm{C_6H_5Br + N_2 + CuCl}$$

  The product of the first step is Bromobenzene.

- **Step (ii):** The bromobenzene is then treated with magnesium in dry ether.

$$C_6H_5Br + Mg \xrightarrow{\text{dry ether}} C_6H_5MgBr$$

This reaction forms the Grignard reagent, phenylmagnesium bromide.

- **Step (iii):** The Grignard reagent is then treated with water. Grignard reagents are strong bases and react with any source of acidic protons, like water.

$$C_6H_5MgBr + H_2O \rightarrow C_6H_6 + Mg(OH)Br$$

The phenyl group acts as a carbanion and abstracts a proton from water, forming benzene (C_6H_6).

The final product of the entire sequence is Benzene.

Step 4: Final Answer:
The final product is Benzene.

Quick Tip

This reaction sequence illustrates important name reactions. The Sandmeyer reaction is a key method to introduce various functional groups onto a benzene ring via a diazonium salt. Remember that Grignard reagents are highly reactive towards protic solvents like water, alcohols, and acids, leading to the formation of the corresponding alkane or arene.

---

57. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R :
Assertion A: Helium is used to dilute oxygen in diving apparatus.
Reason R : Helium has high solubility in O_2.
In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true and R is NOT the correct explanation of A.

Correct Answer: (A) A is true but R is false.

Solution:

Step 1: Understanding the Question:
This is an Assertion-Reason question about the use of helium in diving gas mixtures. We need to evaluate both statements.

Step 3: Detailed Explanation:

Analysis of Assertion A:

The assertion states that helium is used to dilute oxygen in diving apparatus. This is **true**. For deep-sea diving, compressed air is used. However, at high pressures underwater, the nitrogen from the air dissolves in the bloodstream. When the diver ascends, the pressure decreases, and the dissolved nitrogen comes out of solution as bubbles, leading to a painful and dangerous condition called "the bends" or decompression sickness. To avoid this, diving gas mixtures often replace nitrogen with helium. The mixture is called heliox.

Analysis of Reason R:

The reason states that helium has high solubility in O_2. This statement is confusingly worded. The relevant property is not the solubility of helium in oxygen, but the solubility of helium in blood. Helium has very **low** solubility in blood, even at high pressures. This low solubility is the primary reason it is used to replace nitrogen in diving gas. Because very little helium dissolves in the blood, the problem of gas bubbles forming during decompression is minimized. The statement given in the reason, "Helium has high solubility in O_2", is irrelevant and also physically incorrect in the context of explaining its use. The key property is its low solubility in blood. Therefore, the reason is **false**.

Step 4: Final Answer:

Assertion A is a true statement, but Reason R is a false statement.

> **Quick Tip**
>
> The use of helium in diving tanks is a classic application of Henry's Law, which relates the solubility of a gas in a liquid to the partial pressure of the gas above the liquid. Helium's low solubility in blood is the key reason for its use to prevent "the bends".

---

58. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R :

Assertion A: A reaction can have zero activation energy.

Reason R: The minimum extra amount of energy absorbed by reactant molecules so that their energy becomes equal to threshold value, is called activation energy.

In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true and R is NOT the correct explanation of A.

Correct Answer: (B) A is false but R is true.

Solution:

Step 1: Understanding the Question:
This is an Assertion-Reason question. We need to evaluate the correctness of both Assertion A and Reason R and determine if R is the correct explanation for A.

Step 3: Detailed Explanation:
Analysis of Assertion A:
The assertion states that a reaction can have zero activation energy (E_a).
In general, chemical reactions involve the breaking of existing bonds and the formation of new ones. This process requires the reactant molecules to pass through a high-energy transition state. The energy required to reach this transition state from the reactant state is the activation energy.
Therefore, for most chemical reactions, the activation energy is a positive value, meaning an energy barrier must be overcome.
While some specific, very fast reactions (like the recombination of free radicals) are considered to have zero or near-zero activation energy, in the general context of chemical kinetics taught in standard curricula, it's considered that reactions have a non-zero activation energy barrier. Hence, the assertion is generally considered false.

Analysis of Reason R:
The reason provides the definition of activation energy. It states that activation energy is the minimum extra amount of energy that reactant molecules must absorb to reach the threshold energy level (the energy of the transition state).
This is the correct and standard definition of activation energy. Thus, Reason R is a true statement.

Step 4: Final Answer:
Since Assertion A is false and Reason R is true, the correct option is (B).

> **Quick Tip**
>
> In assertion-reason questions, first determine the truth value of each statement independently. If one is true and the other is false, the answer is straightforward. If both are true, you then need to assess if the reason correctly explains the assertion.

---

59. Weight (g) of two moles of the organic compound, which is obtained by heating sodium ethanoate with sodium hydroxide in presence of calcium oxide is :

(A) 30
(B) 18
(C) 16
(D) 32

Correct Answer: (D) 32

Solution:

Step 1: Understanding the Question:
The question asks for the mass of two moles of the organic product formed from a specific chemical reaction: heating sodium ethanoate with soda-lime.

Step 2: Key Formula or Approach:
1. Identify the reaction. Heating a sodium salt of a carboxylic acid with soda-lime (a mixture of NaOH and CaO) is a standard method for the decarboxylation of carboxylic acids, producing an alkane with one less carbon atom. 2. Write the balanced chemical equation. 3. Identify the organic product. 4. Calculate the molar mass of the product. 5. Calculate the mass of two moles of the product.

Step 3: Detailed Explanation:
Part 1: The Reaction
The reaction is the decarboxylation of sodium ethanoate (CH_3COONa) using soda-lime (NaOH + CaO).

$$CH_3COONa + NaOH \xrightarrow{CaO,\Delta} CH_4 + Na_2CO_3$$

The organic product formed is methane (CH_4).

Part 2: Molar Mass of the Product
The molar mass of methane (CH_4) is: Molar Mass $= 12.01$ (C) $+ 4 \times 1.01$ (H) ≈ 16 g/mol.

Part 3: Mass of Two Moles
The question asks for the weight of two moles of the organic compound. Mass = number of moles $\times$ molar mass Mass $= 2$ mol $\times$ 16 g/mol $= 32$ g.

Step 4: Final Answer:
The weight of two moles of the organic compound (methane) is 32 g.

Quick Tip

Decarboxylation with soda-lime is a "step-down" reaction. It removes the -COO group and replaces it with an H atom, effectively reducing the carbon chain length by one. For example, sodium propanoate would yield ethane.

---

60. Amongst the following, the total number of species NOT having eight electrons around central atom in its outer most shell, is
NH_3, $AlCl_3$, $BeCl_2$, CCl_4, PCl_5 :

(A) 4
(B) 1
(C) 3

(D) 2

Correct Answer: (C) 3

Solution:

Step 1: Understanding the Question:
The question asks us to identify and count the number of molecules from a given list that do not follow the octet rule (i.e., the central atom does not have exactly 8 valence electrons).

Step 3: Detailed Explanation:
Let's draw the Lewis structure for each species and count the valence electrons around the central atom.

- **NH_3 (Ammonia):** The central atom is Nitrogen (N). N is in Group 15, so it has 5 valence electrons. It forms 3 single bonds with 3 H atoms and has 1 lone pair. Total electrons around N = 3 (from bonds) $\times$ 2 + 2 (lone pair) = 8 electrons. **(Obeys octet rule)**
- **$AlCl_3$ (Aluminum Chloride):** The central atom is Aluminum (Al). Al is in Group 13, so it has 3 valence electrons. It forms 3 single bonds with 3 Cl atoms. Total electrons around Al = 3 (from bonds) $\times$ 2 = 6 electrons. This is an electron-deficient molecule. **(Does NOT obey octet rule)**
- **$BeCl_2$ (Beryllium Chloride):** The central atom is Beryllium (Be). Be is in Group 2, so it has 2 valence electrons. It forms 2 single bonds with 2 Cl atoms. Total electrons around Be = 2 (from bonds) $\times$ 2 = 4 electrons. This is an electron-deficient molecule. **(Does NOT obey octet rule)**
- **CCl_4 (Carbon Tetrachloride):** The central atom is Carbon (C). C is in Group 14, so it has 4 valence electrons. It forms 4 single bonds with 4 Cl atoms. Total electrons around C = 4 (from bonds) $\times$ 2 = 8 electrons. **(Obeys octet rule)**
- **PCl_5 (Phosphorus Pentachloride):** The central atom is Phosphorus (P). P is in Group 15, so it has 5 valence electrons. It forms 5 single bonds with 5 Cl atoms. Total electrons around P = 5 (from bonds) $\times$ 2 = 10 electrons. This is a hypervalent molecule (expanded octet). **(Does NOT obey octet rule)**

The species that do not have eight electrons around the central atom are $AlCl_3$ (6 electrons), $BeCl_2$ (4 electrons), and PCl_5 (10 electrons).
The total number of such species is 3.

Step 4: Final Answer:
There are 3 species in the list that do not have an octet of electrons around the central atom.

Quick Tip

Exceptions to the octet rule fall into three main categories: 1. **Incomplete Octet:** Central atoms with fewer than 8 electrons (common for elements in Groups 2, 13 like Be, B, Al). 2. **Expanded Octet (Hypervalent):** Central atoms with more than 8 electrons (possible for elements in Period 3 and below, like P, S, Cl). 3. **Odd-Electron Molecules:** Molecules with an odd total number of valence electrons (e.g., NO, ClO_2).

---

61. The relation between n_m, (n_m = the number of permissible values of magnetic quantum number (m)) for a given value of azimuthal quantum number (l), is

(A) $n_m = 2l^2 + 1$
(B) $n_m = l + 2$
(C) $l = \frac{n_m - 1}{2}$
(D) $l = 2n_m + 1$

Correct Answer: (C) $l = \frac{n_m - 1}{2}$

Solution:

Step 1: Understanding the Question:
The question asks for the mathematical relationship between the azimuthal quantum number (l) and the total number of possible values for the magnetic quantum number (m_l), which is denoted here as n_m.

Step 2: Key Formula or Approach:
The rules for quantum numbers state that for a given value of the azimuthal quantum number, l, the magnetic quantum number, m_l, can take any integer value from $-l$ to $+l$, including zero.

$$m_l = -l, -l + 1, ..., 0, ..., l - 1, l$$

Step 3: Detailed Explanation:
To find the total number of permissible values (n_m), we can count the values in the sequence.

- There are l positive values (from 1 to l).
- There are l negative values (from -1 to $-l$).
- There is one zero value.

Total number of values, $n_m = l + l + 1 = 2l + 1$.
So the correct relationship is $n_m = 2l + 1$.
Now, we must check the given options to see which one is equivalent to this relationship.

- (A) $n_m = 2l^2 + 1$: Incorrect.
- (B) $n_m = l + 2$: Incorrect.

- (C) $l = \frac{n_m - 1}{2}$: Let's rearrange this equation to solve for n_m.

$$2l = n_m - 1$$

$$n_m = 2l + 1$$

This matches our derived relationship. So, this option is correct.

- (D) $l = 2n_m + 1$: Incorrect.

Step 4: Final Answer:
The correct relation is given by option (C).

Quick Tip

Remember the dependencies of quantum numbers: n determines the range of l (0 to n-1), and l determines the range of m_l (-l to +l). The number of orbitals in a subshell l is $2l + 1$, and the number of orbitals in a shell n is n^2.

---

62. Which of the following reactions will NOT give primary amine as the product?

(A) $CH_3NC \xrightarrow{\text{(i) LiAlH}_4 \text{ (ii) H}_3\text{O}^+}$ Product
(B) $CH_3CONH_2 \xrightarrow{\text{(i) LiAlH}_4 \text{ (ii) H}_2\text{O}}$ Product
(C) $CH_3CONH_2 \xrightarrow{Br_2/KOH}$ Product
(D) $CH_3CN \xrightarrow{\text{(i) LiAlH}_4 \text{ (ii) H}_3\text{O}^+}$ Product

Correct Answer: (A) $CH_3NC \xrightarrow{\text{(i) LiAlH}_4 \text{ (ii) H}_3\text{O}^+}$ Product

Solution:

Step 1: Understanding the Question:
The question asks to identify which of the given reactions does not produce a primary amine.

Step 2: Key Formula or Approach:
We need to know the products of the reduction of isonitriles, amides, and nitriles, as well as the product of the Hofmann bromamide degradation.

- **Primary amine:** An amine where the nitrogen atom is bonded to one alkyl/aryl group and two hydrogen atoms (R-NH_2).
- **Secondary amine:** An amine where the nitrogen atom is bonded to two alkyl/aryl groups and one hydrogen atom (R_2NH).
- **Tertiary amine:** An amine where the nitrogen atom is bonded to three alkyl/aryl groups (R_3N).

Step 3: Detailed Explanation:
Let's analyze each reaction:

- **(A) Reduction of Methyl isocyanide (CH_3NC):** Isonitriles (or isocyanides) on reduction with $LiAlH_4$ are converted to **secondary amines**.

$$CH_3\text{-}N \equiv C \xrightarrow{LiAlH_4} CH_3\text{-}NH\text{-}CH_3$$

  The product is N-methylmethanamine (dimethylamine), which is a secondary amine. Thus, this reaction does **not** give a primary amine.

- **(B) Reduction of Acetamide (CH_3CONH_2):** Amides are reduced by $LiAlH_4$ to amines. The C=O group is reduced to a CH_2 group.

$$CH_3CONH_2 \xrightarrow{LiAlH_4} CH_3CH_2NH_2$$

  The product is ethanamine, which is a **primary amine**.

- **(C) Hofmann Bromamide Degradation of Acetamide (CH_3CONH_2):** This reaction converts an amide to a primary amine with one less carbon atom.

$$CH_3CONH_2 + Br_2 + 4KOH \rightarrow CH_3NH_2 + 2KBr + K_2CO_3 + 2H_2O$$

  The product is methanamine, which is a **primary amine**.

- **(D) Reduction of Acetonitrile (CH_3CN):** Nitriles (or cyanides) are reduced by $LiAlH_4$ to primary amines. The C≡N group is reduced to a CH_2NH_2 group.

$$CH_3CN \xrightarrow{LiAlH_4} CH_3CH_2NH_2$$

  The product is ethanamine, which is a **primary amine**.

Step 4: Final Answer:
The reduction of methyl isocyanide is the only reaction that does not yield a primary amine; it yields a secondary amine.

Quick Tip

A key distinction to remember for reductions:

- Reduction of Cyanides (R-CN) gives primary amines (R-CH_2NH_2).
- Reduction of Isocyanides (R-NC) gives secondary amines (R-NH-CH_3).

The carbon in the isocyanide group becomes a methyl group attached to the nitrogen.

---

63. Homoleptic complex from the following complexes is :

(A) Pentaamminecarbonatocobalt (III) chloride
(B) Triamminetriaquachromium (III) chloride

(C) Potassium trioxalatoaluminate (III)
(D) Diamminechloridonitrito - N - platinum (II)

Correct Answer: (C) Potassium trioxalatoaluminate (III)

Solution:

Step 1: Understanding the Question:
The question asks us to identify the homoleptic complex among the given options.

Step 2: Key Formula or Approach:
A **homoleptic complex** is a coordination compound in which the central metal ion is coordinated to only one type of ligand.
A **heteroleptic complex** is a coordination compound in which the central metal ion is coordinated to more than one type of ligand.

Step 3: Detailed Explanation:
Let's analyze the ligands in each complex:

- **(A) Pentaamminecarbonatocobalt (III) chloride:** The complex ion is $[Co(NH_3)_5(CO_3)]Cl$. The ligands attached to the cobalt ion are ammine (NH_3) and carbonato ($CO_3{}^{2-}$). Since there are two different types of ligands, this is a heteroleptic complex.
- **(B) Triamminetriaquachromium (III) chloride:** The complex ion is $[Cr(NH_3)_3(H_2O)_3]Cl_3$. The ligands attached to the chromium ion are ammine (NH_3) and aqua (H_2O). Since there are two different types of ligands, this is a heteroleptic complex.
- **(C) Potassium trioxalatoaluminate (III):** The formula is $K_3[Al(C_2O_4)_3]$. The complex ion is $[Al(C_2O_4)_3]^{3-}$. The only ligand attached to the aluminum ion is oxalato ($C_2O_4{}^{2-}$). Since there is only one type of ligand, this is a **homoleptic complex**.
- **(D) Diamminechloridonitrito - N - platinum (II):** The complex is $[Pt(NH_3)_2Cl(NO_2)]$. The ligands attached to the platinum ion are ammine (NH_3), chlorido (Cl^-), and nitrito-N ($NO_2{}^-$). Since there are three different types of ligands, this is a heteroleptic complex.

Step 4: Final Answer:
Potassium trioxalatoaluminate (III) is the only homoleptic complex among the choices.

Quick Tip

Remember "homo" means same and "hetero" means different. In coordination chemistry, "leptic" refers to the ligands. So, homoleptic means "same ligands" and heteroleptic means "different ligands".

---

64. Some tranquilizers are listed below. Which one from the following belongs to barbiturates?

(A) Valium
(B) Veronal
(C) Chlordiazepoxide
(D) Meprobamate

Correct Answer: (B) Veronal

Solution:

Step 1: Understanding the Question:
The question asks us to identify which of the given tranquilizers is a derivative of barbituric acid, i.e., a barbiturate.

Step 3: Detailed Explanation:
Tranquilizers are a class of drugs that reduce stress, anxiety, and tension. They can be broadly classified into different chemical groups. Let's analyze the given options:

- **Valium (Diazepam) and Chlordiazepoxide:** These are well-known tranquilizers belonging to the **benzodiazepine** class of drugs. They are not barbiturates.
- **Meprobamate:** This is also a widely used tranquilizer, but it is classified as a **carbamate**. It is not a barbiturate.
- **Veronal (Barbital):** This is one of the first synthesized barbiturates. Barbiturates are derivatives of barbituric acid and act as central nervous system depressants. Other examples include Luminal (phenobarbital) and Seconal.

Therefore, Veronal is the only barbiturate in the given list.

Step 4: Final Answer:
Veronal is the tranquilizer that belongs to the class of barbiturates.

> **Quick Tip**
>
> It is helpful to memorize the major classes of psychoactive drugs and one or two examples from each class. For tranquilizers, the main classes to know are benzodiazepines (e.g., Valium, Chlordiazepoxide) and barbiturates (e.g., Veronal, Luminal).

---

65. Which amongst the following molecules on polymerization produces neoprene?

(A) $H_2C{=}CH\text{-}C{\equiv}CH$
(B) $CH_3\text{-}C(CH_3){=}CH\text{-}CH_2$
(C) $H_2C{=}CH\text{-}CH{=}CH_2$
(D) $H_2C{=}C(Cl)\text{-}CH{=}CH_2$

Correct Answer: (D) $H_2C=C(Cl)\text{-}CH=CH_2$

Solution:

Step 1: Understanding the Question:
The question asks to identify the monomer unit that polymerizes to form neoprene.

Step 3: Detailed Explanation:
Let's identify the polymers formed from each monomer:

- Neoprene is a synthetic rubber. It is the polymer of the monomer **chloroprene**. The chemical structure of chloroprene is 2-chloro-1,3-butadiene.

$$H_2C=C(Cl)\text{-}CH=CH_2$$

  This matches option (D).

Let's analyze the other options for context:

- Option (A) is vinylacetylene.
- Option (B) is isoprene (2-methyl-1,3-butadiene). Polymerization of isoprene gives polyisoprene, which is the chemical constituent of natural rubber.
- Option (C) is 1,3-butadiene. Polymerization of this monomer gives polybutadiene, another type of synthetic rubber.

Step 4: Final Answer:
The molecule that polymerizes to produce neoprene is chloroprene, which is $H_2C=C(Cl)\text{-}CH=CH_2$.

Quick Tip

Memorize the monomers of important addition polymers:

- Polyethene $\leftarrow$ Ethene
- Teflon $\leftarrow$ Tetrafluoroethene
- PVC $\leftarrow$ Vinyl chloride
- Natural Rubber $\leftarrow$ Isoprene
- Neoprene $\leftarrow$ Chloroprene
- Buna-S $\leftarrow$ 1,3-Butadiene + Styrene
- Buna-N $\leftarrow$ 1,3-Butadiene + Acrylonitrile

---

66. The right option for the mass of CO_2 produced by heating 20 g of 20% pure limestone is (Atomic mass of Ca = 40)

$[\mathbf{CaCO_3 \xrightarrow{1200\ K} CaO + CO_2}]$

(A) 2.64 g
(B) 1.32 g
(C) 1.12 g
(D) 1.76 g

Correct Answer: (D) 1.76 g

Solution:

Step 1: Understanding the Question:
This is a stoichiometry problem. We need to calculate the mass of carbon dioxide (CO_2) produced from the thermal decomposition of an impure sample of limestone (calcium carbonate, $CaCO_3$).

Step 2: Key Formula or Approach:
1. Calculate the mass of the pure reactant from the total mass and purity percentage. 2. Use the balanced chemical equation to establish the molar relationship between the reactant and the product. 3. Convert mass to moles, use the molar ratio, and then convert moles back to mass.

Step 3: Detailed Explanation:
Part 1: Find the mass of pure $CaCO_3$.
Total mass of limestone sample = 20 g.
Purity of limestone = 20 %.
Mass of pure $CaCO_3$ = 20 g $\times \frac{20}{100}$ = 4 g.

Part 2: Use stoichiometry.
The balanced chemical equation is:

$$\mathrm{CaCO_3}(s) \rightarrow \mathrm{CaO}(s) + \mathrm{CO_2}(g)$$

This shows that 1 mole of $CaCO_3$ produces 1 mole of CO_2.

Part 3: Molar mass calculation.
Molar mass of $CaCO_3$ = 40 (Ca) + 12 (C) + 3 × 16 (O) = 100 g/mol.
Molar mass of CO_2 = 12 (C) + 2 × 16 (O) = 44 g/mol.

Part 4: Calculate the mass of CO_2 produced.
From the stoichiometry, 100 g of $CaCO_3$ produces 44 g of CO_2. We have 4 g of pure $CaCO_3$. We can set up a proportion:

$$\frac{\text{mass of } CO_2}{\text{mass of } CaCO_3} = \frac{44 \text{ g}}{100 \text{ g}}$$

$$\text{mass of } CO_2 = \frac{44}{100} \times (\text{mass of } CaCO_3)$$

$$\text{mass of } CO_2 = \frac{44}{100} \times 4 \text{ g} = 0.44 \times 4 \text{ g} = 1.76 \text{ g}$$

Step 4: Final Answer:
The mass of CO_2 produced is 1.76 g.

> Quick Tip
>
> In stoichiometry problems involving purity, always calculate the mass of the pure substance first. The impurities do not participate in the reaction.

---

67. Which one of the following statements is correct?

(A) The bone in human body is an inert and unchanging substance.
(B) Mg plays roles in neuromuscular function and interneuronal transmission.
(C) The daily requirement of Mg and Ca in the human body is estimated to be 0.2 - 0.3 g.
(D) All enzymes that utilise ATP in phosphate transfer require Ca as the cofactor.

Correct Answer: (B) Mg plays roles in neuromuscular function and interneuronal transmission.

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statement among four options related to the biological roles of calcium and magnesium.

Step 3: Detailed Explanation:
Let's analyze each statement:

- **(A) The bone in human body is an inert and unchanging substance.**
  This is **false**. Bone is a dynamic living tissue that is constantly being remodeled (broken down and rebuilt) throughout life. It is not inert.
- **(B) Mg plays roles in neuromuscular function and interneuronal transmission.**
  This is **true**. Magnesium ions (Mg^{2+}) are crucial for many biological processes. They act as a physiological calcium channel blocker and are essential for nerve impulse transmission, muscle contraction (neuromuscular function), and maintaining a normal heart rhythm.
- **(C) The daily requirement of Mg and Ca in the human body is estimated to be 0.2 - 0.3 g.**
  This is **false**. The daily requirement for an adult for calcium is much higher, typically around 1000-1200 mg (1.0-1.2 g). For magnesium, it is around 300-400 mg (0.3-0.4 g). The given range of 200-300 mg (0.2-0.3 g) is too low, especially for calcium.
- **(D) All enzymes that utilise ATP in phosphate transfer require Ca as the cofactor.**
  This is **false**. Most enzymes that utilize ATP in phosphate transfer (kinases) require **Magnesium (Mg^{2+})** as the cofactor. The Mg^{2+} ion complexes with the phosphate

groups of ATP, stabilizing the negative charges and facilitating the nucleophilic attack for phosphate transfer.

Step 4: Final Answer:
The only correct statement is (B).

> Quick Tip
>
> Remember the key biological roles: Calcium (Ca^{2+}) is primarily associated with bones, teeth, muscle contraction signaling, and blood clotting. Magnesium (Mg^{2+}) is a critical cofactor for ATP-dependent enzymes and plays a key role in nerve and muscle function.

---

68. Match List - I with List - II:

List - I	List - II
A. Coke	**I. Carbon atoms are sp^3 hybridised.**
B. Diamond	**II. Used as a dry lubricant**
C. Fullerene	**III. Used as a reducing agent**
D. Graphite	**IV. Cage like molecules**

Choose the correct answer from the options given below :

(A) A-III, B-I, C-IV, D-II
(B) A-III, B-IV, C-I, D-II
(C) A-II, B-IV, C-I, D-III
(D) A-IV, B-I, C-II, D-III

Correct Answer: (A) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:
This is a matching question where we need to correctly pair different forms of carbon (allotropes and an amorphous form) with their characteristic properties or uses.

Step 3: Detailed Explanation:
Let's match each item in List-I with its correct description in List-II.

- **A. Coke:** Coke is an amorphous form of carbon produced by heating coal in the absence of air. It is a key material in metallurgy, where it acts as a **reducing agent** to reduce metal oxides to metals (e.g., in a blast furnace). So, **A matches with III**.
- **B. Diamond:** Diamond is a crystalline allotrope of carbon. In its structure, each carbon atom is bonded to four other carbon atoms in a tetrahedral arrangement. This corresponds to **sp^3 hybridization**. So, **B matches with I**.
- **C. Fullerene:** Fullerenes (like C_60) are crystalline allotropes of carbon where the atoms are arranged in a hollow sphere or ellipsoid. These are described as **cage-like molecules** (often called buckyballs). So, **C matches with IV**.

- **D. Graphite:** Graphite is a crystalline allotrope of carbon with a layered structure. The layers can slide easily over one another, which makes graphite soft and an excellent **dry lubricant**. So, **D matches with II**.

The correct matching is: A-III, B-I, C-IV, D-II.

Step 4: Final Answer:
Comparing our matching with the options, option (A) is the correct one.

> Quick Tip
>
> Associate the structure of carbon allotropes with their properties:
>
> - **Diamond:** 3D network, sp^3, hard.
> - **Graphite:** 2D layers, sp^2, soft, lubricant, conductive.
> - **Fullerene:** Cage-like, sp^2, buckyballs.
> - **Coke/Charcoal:** Amorphous, porous, good reducing agent/adsorbent.

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69. The stability of Cu^{2+} is more than Cu^{+} salts in aqueous solution due to -

(A) hydration energy.
(B) second ionisation enthalpy.
(C) first ionisation enthalpy.
(D) enthalpy of atomization.

Correct Answer: (A) hydration energy.

Solution:

Step 1: Understanding the Question:
The question asks for the reason why the Cu^{2+} ion is more stable than the Cu^{+} ion specifically in aqueous solutions, despite the fact that removing a second electron from copper requires more energy.

Step 3: Detailed Explanation:
Let's analyze the factors involved:

- **Ionization Enthalpy:** The electronic configuration of Cu is [Ar] $3d^{10}$ $4s^1$.
  - The first ionization enthalpy (IE_1) to form Cu^+ ([Ar] $3d^{10}$) is relatively low.
  - The second ionization enthalpy (IE_2) to form Cu^{2+} ([Ar] $3d^9$) is very high because it involves removing an electron from a very stable, completely filled d-orbital.

Based on ionization enthalpies alone, Cu^+ should be much more stable than Cu^{2+}. This contradicts the observation in aqueous solution. Therefore, (B) and (C) are not the reasons for the stability of Cu^{2+}; in fact, they argue against it.

- **Hydration Energy (Enthalpy):** When ions are dissolved in water, they become hydrated, and this process releases energy. This released energy is called hydration energy. The magnitude of hydration energy depends on the charge density of the ion (charge/size ratio).
  - The Cu^{2+} ion has a greater positive charge (+2) and a smaller ionic radius compared to the Cu^+ ion (+1).
  - This gives Cu^{2+} a much higher charge density than Cu^+.
  - Consequently, the hydration enthalpy of Cu^{2+} is much more negative (i.e., much more energy is released) than that of Cu^+.
- **Overall Stability:** The high amount of energy released during the hydration of Cu^{2+} more than compensates for the high second ionization enthalpy required to form it. This large negative hydration enthalpy makes the overall process of forming hydrated Cu^{2+} from Cu(s) more favorable than forming hydrated Cu^+, thus making Cu^{2+} more stable in aqueous solutions.
- **Enthalpy of atomization** is the energy required to convert a substance from its standard state to gaseous atoms. While part of the overall Born-Haber cycle, it is not the primary differentiating factor for the stability of the two ions in solution.

Step 4: Final Answer:
The greater stability of Cu^{2+} in aqueous solution is due to its much higher hydration energy.

Quick Tip

Stability of ions can be different in the gaseous phase versus aqueous solution. In the gas phase, stability is governed mainly by ionization energy (favoring Cu^+). In aqueous solution, the hydration energy term becomes dominant and can reverse the stability order (favoring Cu^{2+}).

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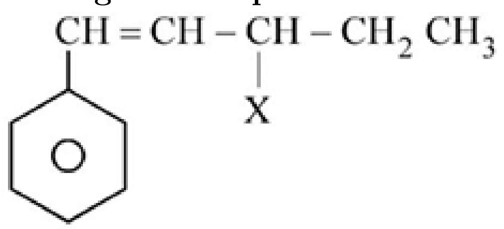
70. The given compound is an example of



(A) allylic halide
(B) vinylic halide
(C) benzylic halide

(D) aryl halide

Correct Answer: (A) allylic halide

Solution:

Step 1: Understanding the Question:
The question asks to classify the given organic halide based on the position of the halogen atom (X) relative to a carbon-carbon double bond.

Step 3: Detailed Explanation:
Let's define the different types of halides mentioned:

- **Vinylic halide:** The halogen atom is bonded directly to one of the sp^2-hybridized carbon atoms of a C=C double bond.
- **Allylic halide:** The halogen atom is bonded to an sp^3-hybridized carbon atom which is adjacent to a C=C double bond. This carbon is called the allylic carbon.
- **Benzylic halide:** The halogen atom is bonded to an sp^3-hybridized carbon atom which is directly attached to a benzene ring.
- **Aryl halide:** The halogen atom is bonded directly to an sp^2-hybridized carbon atom of a benzene ring.

Now let's analyze the given structure: $CH_2{=}CH\text{-}\mathbf{C}H(X)\text{-}CH_2CH_3$.

- The halogen atom X is attached to the third carbon atom in the chain.
- This carbon atom is sp^3-hybridized (it forms four single bonds).
- This sp^3-hybridized carbon is directly attached to the sp^2-hybridized carbon of the $CH_2{=}CH$- double bond.

This fits the definition of an allylic halide perfectly.

Step 4: Final Answer:
The given compound is an example of an allylic halide.

> Quick Tip
>
> To quickly classify halides, look at the carbon atom attached to the halogen. If it's sp^3 and next to a C=C bond, it's allylic. If it's sp^2 and part of a C=C bond, it's vinylic. If it's sp^3 and next to a benzene ring, it's benzylic. If it's sp^2 and part of a benzene ring, it's aryl.

---

71. Given below are two statements :
Statement I: A unit formed by the attachment of a base to 1' position of sugar is

known as nucleoside
Statement II: When nucleoside is linked to phosphorous acid at 5'-position of sugar moiety, we get nucleotide.
In the light of the above statements, choose the correct answer from the options given below :

(A) Statement I is true but Statement II is false.
(B) Statement I is false but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (A) Statement I is true but Statement II is false.

Solution:

Step 1: Understanding the Question:
The question asks us to evaluate two statements regarding the definitions of nucleosides and nucleotides.

Step 3: Detailed Explanation:
Analysis of Statement I:
A nucleoside is a structural subunit of nucleic acids (DNA and RNA). It consists of two components: a nitrogenous base (a purine or a pyrimidine) and a five-carbon sugar (ribose in RNA or deoxyribose in DNA). The nitrogenous base is attached to the 1' carbon of the sugar via an N-glycosidic bond. This statement accurately defines a nucleoside. Therefore, Statement I is **true**.

Analysis of Statement II:
A nucleotide is formed when a phosphate group is attached to a nucleoside. The phosphate group is linked to the sugar moiety, typically at the 5' position, via a phosphoester bond. The statement says that the nucleoside is linked to **phosphorous acid**. This is incorrect. The phosphate group in a nucleotide is derived from **phosphoric acid** (H_3PO_4), not phosphorous acid (H_3PO_3). Therefore, Statement II is **false**.

Step 4: Final Answer:
Statement I is true, but Statement II is false.

Quick Tip

Remember the building blocks of nucleic acids:

- Base + Sugar = Nucleo**s**ide
- Base + Sugar + Phosphate = Nucleo**t**ide
- (The 't' in nucleotide can remind you of the 'three' components).

Also, be mindful of the difference between phosphoric acid (H_3PO_4) and phosphorous acid (H_3PO_3).

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72. In Lassaigne's extract of an organic compound, both nitrogen and sulphur are present, which gives blood red colour with Fe^{3+} due to the formation of -

(A) $[Fe(CN)_5NOS]^{4-}$
(B) $[Fe(SCN)]^{2+}$
(C) $Fe_4[Fe(CN)_6]_3 \cdot xH_2O$
(D) NaSCN

Correct Answer: (B) $[Fe(SCN)]^{2+}$

Solution:

Step 1: Understanding the Question:
This question is about the qualitative analysis of organic compounds, specifically the Lassaigne's test for nitrogen and sulphur when present together. We need to identify the chemical species responsible for the blood-red color.

Step 2: Key Formula or Approach:
In Lassaigne's test, the organic compound is fused with sodium metal.

- If nitrogen is present, sodium cyanide (NaCN) is formed.
- If sulphur is present, sodium sulfide (Na_2S) is formed.
- If both nitrogen and sulphur are present, sodium thiocyanate (NaSCN) is formed.

$$\mathrm{Na + C + N + S \rightarrow NaSCN}$$

The test for thiocyanate involves adding a neutral or slightly acidic solution of iron(III) chloride ($FeCl_3$).

Step 3: Detailed Explanation:
When the Lassaigne's extract containing sodium thiocyanate (NaSCN) is treated with Fe^{3+} ions, a complex is formed. The thiocyanate ion (SCN^-) reacts with the ferric ion (Fe^{3+}) to produce a range of iron(III) thiocyanate complexes, which are intensely blood-red in color.

$$\mathrm{Fe^{3+} + SCN^- \rightleftharpoons [Fe(SCN)]^{2+}}$$

Further reactions can form $[Fe(SCN)_2]^+$, $[Fe(SCN)_3]$, etc., all of which contribute to the blood-red coloration. The simplest and most commonly written product is $[Fe(SCN)]^{2+}$. Let's analyze the options:

- (A) $[Fe(CN)_5NOS]^{4-}$: This is the sodium nitroprusside complex, which gives a violet color with sulfide ions (S^{2-}), not a blood-red color with Fe^{3+}.
- (B) $[Fe(SCN)]^{2+}$: This is the iron(III) thiocyanate complex, which is responsible for the characteristic blood-red color. This is the correct answer.
- (C) $Fe_4[Fe(CN)_6]_3 \cdot xH_2O$: This is Prussian blue, formed when nitrogen alone is present (as NaCN) and treated with $FeSO_4$ followed by $FeCl_3$.
- (D) NaSCN: This is the substance formed during the sodium fusion, but it is colorless. The color appears only after adding Fe^{3+}.

Step 4: Final Answer:
The formation of the complex ion $[Fe(SCN)]^{2+}$ is responsible for the blood-red color.

Quick Tip

Remember the key colorimetric tests in Lassaigne's analysis:

- N only: Prussian blue with $FeSO_4/FeCl_3$.
- S only: Violet color with sodium nitroprusside.
- N and S together: Blood-red color with $FeCl_3$.
- Halogens: White/Yellow precipitate with $AgNO_3$.

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73. The number of σ bonds, π bonds and lone pair of electrons in pyridine, respectively are:

(A) 11, 3, 1
(B) 12, 2, 1
(C) 11, 2, 0
(D) 12, 3, 0

Correct Answer: (A) 11, 3, 1

Solution:

Step 1: Understanding the Question:
We need to count the number of sigma (σ) bonds, pi (π) bonds, and lone pairs of electrons in the structure of pyridine.

Step 2: Key Formula or Approach:
1. Draw the Lewis structure of pyridine (C_5H_5N). 2. Count the bonds and lone pairs based on the structure.

- Every single bond is one σ bond.
- Every double bond consists of one σ bond and one π bond.
- Every triple bond consists of one σ bond and two π bonds.
- Determine the lone pairs on the heteroatom (Nitrogen) based on its valency.

Step 3: Detailed Explanation:
The structure of pyridine is a six-membered aromatic ring containing five carbon atoms and one nitrogen atom. Each carbon is bonded to one hydrogen atom.
Let's count the bonds:

- **σ bonds:**
  - There are 5 C-H single bonds. (5 σ bonds)
  - Within the ring, there are 4 C-C single bonds and 2 C-N single bonds. (6 σ bonds)
  - Total σ bonds = 5 (C-H) + 6 (in-ring single bonds) = 11 σ bonds.

  Alternatively, for a cyclic molecule, the number of σ bonds equals the number of atoms. Pyridine has 5 C + 5 H + 1 N = 11 atoms, so there are 11 σ bonds. This shortcut works for single rings.
- **π bonds:**
  - Pyridine has three double bonds in the ring to satisfy aromaticity. Each double bond contains one π bond.
  - Total π bonds = 3.
- **Lone pairs:**
  - Carbon has 4 valence electrons and forms 4 bonds, so no lone pairs on carbon.
  - Nitrogen is in group 15 and has 5 valence electrons. In pyridine, it forms three bonds (two single, one double). So, the number of non-bonding electrons = 5 - 3 = 2.
  - This corresponds to one lone pair of electrons.

So, the counts are: 11 σ bonds, 3 π bonds, and 1 lone pair.

Step 4: Final Answer:
The correct combination is 11, 3, 1.

Quick Tip

For polyatomic molecules, a quick way to count σ bonds is: (Total number of atoms) - 1 for acyclic molecules, or (Total number of atoms) for monocyclic molecules. The number of π bonds is easily counted from the double and triple bonds in the structure.

74. A compound is formed by two elements A and B. The element B forms cubic close packed structure and atoms of A occupy 1/3 of tetrahedral voids. If the formula of the compound is A_xB_y, then the value of x + y is in option

(A) 3
(B) 2
(C) 5
(D) 4

Correct Answer: (C) 5

Solution:

Step 1: Understanding the Question:
This is a solid-state chemistry problem. We need to determine the empirical formula of a compound based on the crystal lattice structure formed by its constituent elements.

Step 2: Key Formula or Approach:
1. In a close-packed structure (like CCP or FCC), if there are N atoms forming the lattice, then:

- The number of octahedral voids is N.
- The number of tetrahedral voids is 2N.

2. The cubic close-packed (CCP) structure is the same as the face-centered cubic (FCC) structure. The effective number of atoms per unit cell for CCP is N=4. We can work with a general N and the ratio will be the same. 3. Determine the number of atoms of A and B per unit cell based on the information given. 4. Find the simplest whole-number ratio of A to B to get the formula A_xB_y.

Step 3: Detailed Explanation:
Let's assume the number of atoms of element B forming the CCP lattice is N.

- Number of B atoms = N.

The number of tetrahedral voids in this lattice is 2N.
Atoms of element A occupy 1/3 of these tetrahedral voids.

- Number of A atoms = $\frac{1}{3} \times$ (Number of tetrahedral voids) = $\frac{1}{3} \times (2N) = \frac{2N}{3}$.

Now, we find the ratio of the number of atoms of A to B:

$$\text{Ratio A : B} = \frac{2N}{3} : N$$

To get a simple ratio, we can divide by N:

$$\text{Ratio A : B} = \frac{2}{3} : 1$$

To express this in whole numbers, we multiply by 3:

$$\text{Ratio A : B} = 2 : 3$$

So, the empirical formula of the compound is A_2B_3.
The question gives the formula as A_xB_y. By comparing, we have $x = 2$ and $y = 3$.
We need to find the value of $x + y$.

$$x + y = 2 + 3 = 5$$

Step 4: Final Answer:
The value of x + y is 5.

Quick Tip

In lattice problems, it's often easiest to assume the number of lattice-forming atoms (N) is 1 and work with fractions. - B atoms = 1 - Tetrahedral voids = 2 - A atoms = (1/3) * 2 = 2/3 - Ratio A:B = 2/3 : 1 = 2:3. This method avoids carrying 'N' through the calculation.

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75. The correct order of energies of molecular orbitals of N_2 molecule, is :

(A) $\sigma 1s < \sigma^*1s < \sigma 2s < \sigma^*2s < \sigma 2p_z < (\pi 2p_x = \pi 2p_y) < (\pi^*2p_x = \pi^*2p_y) < \sigma^*2p_z$
(B) $\sigma 1s < \sigma^*1s < \sigma 2s < \sigma^*2s < (\pi 2p_x = \pi 2p_y) < (\pi^*2p_x = \pi^*2p_y) < \sigma 2p_z < \sigma^*2p_z$
(C) $\sigma 1s < \sigma^*1s < \sigma 2s < \sigma^*2s < (\pi 2p_x = \pi 2p_y) < \sigma 2p_z < (\pi^*2p_x = \pi^*2p_y) < \sigma^*2p_z$
(D) $\sigma 1s < \sigma^*1s < \sigma 2s < \sigma^*2s < \sigma 2p_z < (\pi 2p_x = \pi 2p_y) < (\pi^*2p_x = \pi^*2p_y) < \sigma^*2p_z$

Correct Answer: (C) $\sigma 1s < \sigma^*1s < \sigma 2s < \sigma^*2s < (\pi 2p_x = \pi 2p_y) < \sigma 2p_z < (\pi^*2p_x = \pi^*2p_y) < \sigma^*2p_z$

Solution:

Step 1: Understanding the Question:
The question asks for the correct sequence of molecular orbitals (MOs) in increasing order of their energy for the dinitrogen (N_2) molecule.

Step 2: Key Formula or Approach:
According to Molecular Orbital Theory (MOT), the energy order of MOs for diatomic molecules depends on the extent of s-p orbital mixing.
For lighter diatomic molecules like B_2, C_2, and N_2 (up to 14 electrons), significant s-p mixing occurs. This mixing raises the energy of the $\sigma 2p_z$ orbital above that of the $\pi 2p_x$ and $\pi 2p_y$ orbitals.
For heavier diatomic molecules like O_2 and F_2, s-p mixing is less significant, and the $\sigma 2p_z$ orbital has lower energy than the π orbitals.

Step 3: Detailed Explanation:
The N_2 molecule has a total of 14 electrons. Due to s-p mixing, the correct energy order of its molecular orbitals is:

$$\sigma 1s < \sigma^* 1s < \sigma 2s < \sigma^* 2s < (\pi 2p_x = \pi 2p_y) < \sigma 2p_z < (\pi^* 2p_x = \pi^* 2p_y) < \sigma^* 2p_z$$

Let's compare this correct order with the given options.

- Option (A) places $\sigma 2p_z$ before $(\pi 2p_x = \pi 2p_y)$. This is incorrect for N_2.
- Option (B) has an unusual order with antibonding π^* orbitals before the bonding $\sigma 2p_z$. This is incorrect.
- Option (C) matches the correct energy order for N_2 derived from the principle of s-p mixing.
- Option (D) is identical to option (A) and is incorrect. (Note: There might be a typo in the original question options, but (C) is the only one that represents the correct order for N_2).

Step 4: Final Answer:
The correct order of energies of molecular orbitals of the N_2 molecule is given in option (C).

Quick Tip

A simple way to remember the MO order is to check the total number of electrons. For diatomic molecules with ≤ 14 electrons, the $\pi 2p$ orbitals are filled before the $\sigma 2p_z$ orbital. For molecules with > 14 electrons, the $\sigma 2p_z$ is filled first.

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76. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R :
Assertion A: Metallic sodium dissolves in liquid ammonia giving a deep blue solution, which is paramagnetic.
Reason R: The deep blue solution is due to the formation of amide.
In the light of the above statements, choose the correct answer from the options given below:

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (A) A is true but R is false.

Solution:

Step 1: Understanding the Question:
This Assertion-Reason question deals with the properties of solutions of alkali metals in liquid ammonia. We need to evaluate both statements.

Step 3: Detailed Explanation:
Analysis of Assertion A:
The assertion states that sodium metal dissolves in liquid ammonia to form a deep blue, paramagnetic solution. This is a well-known characteristic property of alkali metals.

$$Na(s) + (x+y)NH_3(l) \rightarrow [Na(NH_3)_x]^+ + [e(NH_3)_y]^-$$

The dissolution produces solvated cations (ammoniated cations) and solvated electrons (ammoniated electrons).

- The presence of unpaired ammoniated electrons makes the solution **paramagnetic**.
- These ammoniated electrons absorb energy in the visible region of the electromagnetic spectrum, and the transmitted light is deep blue.

Therefore, Assertion A is **true**.

Analysis of Reason R:
The reason claims that the deep blue color is due to the formation of amide. Sodium amide ($NaNH_2$) is formed when the solution is allowed to stand for a long time or in the presence of a catalyst (like Fe^{3+} ions), as the ammoniated electron slowly reduces ammonia:

$$[e(NH_3)_y]^- + NH_3 \rightarrow NH_2^- + \frac{1}{2}H_2 + (y-1)NH_3$$

This is a decomposition reaction that causes the blue color to fade. The amide ion (NH_2^-) itself is not responsible for the initial deep blue color. The color is due to the ammoniated electrons. Therefore, Reason R is **false**.

Step 4: Final Answer:
Assertion A is true, but Reason R is false.

Quick Tip

Remember the key species in alkali metal-ammonia solutions:

- **Dilute solutions:** Blue color and paramagnetism due to ammoniated electrons.
- **Concentrated solutions:** Bronze color and diamagnetism due to the formation of electron pairs and metal clusters.
- **Decomposition product:** Sodium amide ($NaNH_2$), which is colorless.

---

77. Which one is an example of heterogenous catalysis?

(A) Decomposition of ozone in presence of nitrogen monoxide.
(B) Combination between dinitrogen and dihydrogen to form ammonia in the presence of finely divided iron.
(C) Oxidation of sulphur dioxide into sulphur trioxide in the presence of oxides of nitrogen.
(D) Hydrolysis of sugar catalysed by H^+ ions.

Correct Answer: (B) Combination between dinitrogen and dihydrogen to form ammonia in the presence of finely divided iron.

Solution:

Step 1: Understanding the Question:
The question asks us to identify an example of heterogeneous catalysis from the given options.

Step 2: Key Formula or Approach:
Catalysis is classified based on the physical state (phase) of the reactants and the catalyst.

- **Homogeneous catalysis:** The reactants and the catalyst are in the same phase (e.g., all are gases, or all are in the same liquid solution).
- **Heterogeneous catalysis:** The reactants and the catalyst are in different phases (e.g., gaseous reactants with a solid catalyst).

Step 3: Detailed Explanation:
Let's analyze the phase of reactants and catalysts in each option:

- **(A) Decomposition of ozone in presence of nitrogen monoxide.**
  Reaction: $2O_3(g) \xrightarrow{NO(g)} 3O_2(g)$.
  Reactant (Ozone) is a gas, and the catalyst (NO) is also a gas. Since they are in the same phase, this is **homogeneous catalysis**.
- **(B) Combination between dinitrogen and dihydrogen to form ammonia in the presence of finely divided iron.**
  This is the Haber-Bosch process: $N_2(g) + 3H_2(g) \xrightarrow{Fe(s)} 2NH_3(g)$.
  Reactants (N_2, H_2) are gases, but the catalyst (Iron) is a solid. Since the reactants and catalyst are in different phases, this is **heterogeneous catalysis**.
- **(C) Oxidation of sulphur dioxide into sulphur trioxide in the presence of oxides of nitrogen.**
  This is the lead chamber process: $2SO_2(g) + O_2(g) \xrightarrow{NO(g)} 2SO_3(g)$.
  Reactants (SO_2, O_2) are gases, and the catalyst (NO) is also a gas. This is **homogeneous catalysis**.
- **(D) Hydrolysis of sugar catalysed by H^+ ions.**
  Reaction: $C_{12}H_{22}O_{11}(aq) + H_2O(l) \xrightarrow{H^+(aq)} C_6H_{12}O_6(aq) + C_6H_{12}O_6(aq)$.
  The reactant (sugar) and the catalyst (H^+ ions) are both dissolved in the same aqueous solution. This is **homogeneous catalysis**.

Step 4: Final Answer:
The Haber-Bosch process for ammonia synthesis is the correct example of heterogeneous catalysis.

Quick Tip

Heterogeneous catalysis often involves a solid catalyst and gaseous or liquid reactants. This is very common in industrial processes (like the Haber-Bosch process or catalytic converters in cars) because it makes separating the catalyst from the products easier.

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78. For a certain reaction, the rate = k[A]^2[B], when the initial concentration of A is tripled keeping concentration of B constant, the initial rate would

(A) increase by a factor of nine.
(B) increase by a factor of three.
(C) decrease by a factor of nine.
(D) increase by a factor of six.

Correct Answer: (A) increase by a factor of nine.

Solution:

Step 1: Understanding the Question:
This is a chemical kinetics problem. We are given a rate law and asked to determine how the reaction rate changes when the concentration of one of the reactants is changed.

Step 2: Key Formula or Approach:
The rate law is given as: Rate $= k[A]^2[B]$. We need to compare the initial rate with the new rate after changing the concentration of A. Let the initial rate be Rate$_1$.

$$\text{Rate}_1 = k[A]^2[B]$$

Let the new rate be Rate$_2$, where the concentration of A is tripled. The new concentration of A, [A]$_{new}$, is 3[A]. The concentration of B remains constant.

$$\text{Rate}_2 = k[A]_{new}^2[B] = k(3[A])^2[B]$$

Step 3: Detailed Explanation:
Let's calculate the new rate, Rate$_2$:

$$\text{Rate}_2 = k(3[A])^2[B] = k(9[A]^2)[B] = 9 \times (k[A]^2[B])$$

We know that Rate$_1 = k[A]^2[B]$. So, we can substitute this into the equation for Rate$_2$:

$$\text{Rate}_2 = 9 \times \text{Rate}_1$$

This shows that the new rate is nine times the initial rate. The rate increases by a factor of nine.

Step 4: Final Answer:
The initial rate would increase by a factor of nine.

Quick Tip

To find the effect of changing a reactant's concentration, simply look at its order in the rate law. If the concentration of a reactant is changed by a factor 'x' and its order is 'n', the rate will change by a factor of x^n. In this case, [A] is tripled (x=3) and its order is 2 (n=2), so the rate changes by a factor of $3^2 = 9$.

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79. Select the correct statements from the following:
A. Atoms of all elements are composed of two fundamental particles.
B. The mass of the electron is 9.10939 $\times$ 10^{-31} kg.
C. All the isotopes of a given element show same chemical properties.
D. Protons and electrons are collectively known as nucleons.
E. Dalton's atomic theory, regarded the atom as an ultimate particle of matter.
Choose the correct answer from the options given below :

(A) A and E only
(B) B, C and E only
(C) A, B and C only
(D) C, D and E only

Correct Answer: (B) B, C and E only

Solution:

Step 1: Understanding the Question:
We need to evaluate a list of five statements related to atomic structure and theory and identify the combination of correct statements.

Step 3: Detailed Explanation:
Let's analyze each statement:

- **Statement A: Atoms of all elements are composed of two fundamental particles.**
This is **false**. Atoms are generally composed of three fundamental particles: protons, neutrons, and electrons. An exception is the protium isotope of hydrogen (^1H), which has one proton and one electron but no neutron. However, the statement claims this for "all elements," which is incorrect.
- **Statement B: The mass of the electron is 9.10939 $\times$ 10^{-31} kg.**
This is **true**. This is the experimentally determined and accepted value for the rest mass of an electron.

- **Statement C: All the isotopes of a given element show same chemical properties.**
This is **true**. Isotopes are atoms of the same element with the same number of protons but different numbers of neutrons. Since they have the same number of protons, they also have the same number of electrons in their neutral state. Chemical properties are primarily determined by the electronic configuration, which is the same for all isotopes of an element.
- **Statement D: Protons and electrons are collectively known as nucleons.**
This is **false**. The term "nucleons" refers to the particles found in the atomic nucleus, which are protons and neutrons. Electrons are not nucleons; they orbit the nucleus.
- **Statement E: Dalton's atomic theory, regarded the atom as an ultimate particle of matter.**
This is **true**. One of the key postulates of John Dalton's original atomic theory (early 19th century) was that atoms are fundamental, indivisible particles. Although we now know atoms are divisible, the statement correctly describes a core aspect of Dalton's theory.

So, the correct statements are B, C, and E.

Step 4: Final Answer:
The correct option is (B), which includes statements B, C, and E.

Quick Tip

When evaluating historical scientific theories like Dalton's, answer based on the theory's original postulates, not our modern understanding. The statement about Dalton's theory is correct in its historical context.

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80. Consider the following reaction and identify the product (P).
CH_3-$CH(CH_3)$-$CH(OH)$-CH_3 + HBr $\rightarrow$ Product (P)
(3-Methylbutan-2-ol)

(A) CH_3-$CH(CH_3)$-$CH(Br)$-CH_3
(B) CH_3-$C(CH_3)(Br)$-CH_2-CH_3
(C) CH_3-$C(Br)(CH_3)$-CH_2-CH_3
(D) CH_3-$CH{=}CH$-CH_3

Correct Answer: (C) CH_3-$C(Br)(CH_3)$-CH_2-CH_3

Solution:

Step 1: Understanding the Question:
This is a reaction of a secondary alcohol with hydrogen bromide (HBr). We need to predict the major product, which involves understanding the reaction mechanism.

Step 2: Key Formula or Approach:
The reaction of an alcohol with a hydrogen halide (like HBr) proceeds via a carbocation intermediate (for secondary and tertiary alcohols, following an S_N1 mechanism). The mechanism involves three steps: 1. Protonation of the alcohol's hydroxyl group by H^+ to form a good leaving group (water). 2. Loss of the leaving group (H_2O) to form a carbocation. 3. Attack of the nucleophile (Br^-) on the carbocation. A key feature of this mechanism is the possibility of carbocation rearrangement to form a more stable carbocation. The stability of carbocations follows the order: tertiary ¿ secondary ¿ primary.

Step 3: Detailed Explanation:
Step 1: Protonation
The oxygen atom of the -OH group in 3-methylbutan-2-ol gets protonated by HBr.

$$\text{CH}_3\text{-CH(CH}_3\text{)-CH(OH)-CH}_3 + \text{H}^+ \rightarrow \text{CH}_3\text{-CH(CH}_3\text{)-CH(OH}_2^+\text{)-CH}_3$$

Step 2: Formation of Carbocation
The protonated alcohol loses a water molecule to form a secondary carbocation.

$$\text{CH}_3\text{-CH(CH}_3\text{)-CH(OH}_2^+\text{)-CH}_3 \rightarrow \text{CH}_3\text{-CH(CH}_3\text{)-CH}^+\text{-CH}_3 + \text{H}_2\text{O}$$

This is a secondary (2°) carbocation.
Step 3: Carbocation Rearrangement
We must check if this carbocation can rearrange to a more stable one. A hydride shift (1,2-hydride shift) from the adjacent carbon (carbon-3) can occur.

$$\text{CH}_3\text{-}\overset{\text{H}}{\overset{|}{\text{C}}}\text{(CH}_3\text{)-}\overset{+}{\text{C}}\text{H-CH}_3 \xrightarrow{\text{1,2-hydride shift}} \text{CH}_3\text{-}\overset{+}{\text{C}}\text{(CH}_3\text{)-CH}_2\text{-CH}_3$$

The new carbocation formed is a tertiary (3°) carbocation, which is significantly more stable than the initial secondary carbocation. This rearrangement will be the major pathway.
Step 4: Nucleophilic Attack
The bromide ion (Br^-) now attacks the more stable tertiary carbocation.

$$\text{CH}_3\text{-}\overset{+}{\text{C}}\text{(CH}_3\text{)-CH}_2\text{-CH}_3 + \text{Br}^- \rightarrow \text{CH}_3\text{-C(Br)(CH}_3\text{)-CH}_2\text{-CH}_3$$

The major product is 2-bromo-2-methylbutane. This matches the structure in option (C). The product in option (A) would be formed if no rearrangement occurred, making it the minor product. Option (D) is an elimination product, which is less likely under these conditions.

Step 4: Final Answer:
The major product (P) is 2-bromo-2-methylbutane, formed via a carbocation rearrangement.

Quick Tip

Whenever a reaction proceeds through a carbocation intermediate (like S_N1 reactions of alcohols or additions to alkenes), always check for the possibility of rearrangement (hydride or alkyl shift) to form a more stable carbocation. This is a very common topic for questions.

81. Taking stability as the factor, which one of the following represents correct relationship?

(A) $AlCl > AlCl_3$
(B) $TlI > TlI_3$
(C) $TlCl_3 > TlCl$
(D) $InI_3 > InI$

Correct Answer: (B) $TlI > TlI_3$

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct stability relationship for halides of Group 13 elements. This relates to the concept of the inert pair effect.

Step 2: Key Formula or Approach:
The **inert pair effect** describes the increasing stability of the lower oxidation state (n-2) compared to the higher group oxidation state (n) as we move down a p-block group. For Group 13 (B, Al, Ga, In, Tl), the group oxidation state is +3, and the lower oxidation state is +1.
The inert pair effect becomes significant for the heavier elements in the group. This means that for Thallium (Tl), the +1 oxidation state is more stable than the +3 oxidation state. For the lighter elements like Aluminum (Al), the +3 oxidation state is the most stable.

Step 3: Detailed Explanation:
Let's analyze the stability of the oxidation states for the elements in the options:

- **Aluminum (Al):** It is in the 3rd period. The inert pair effect is negligible. The +3 oxidation state is much more stable than the +1 state. Therefore, $AlCl_3$ is much more stable than AlCl. The relationship AlCl ¿ $AlCl_3$ is **incorrect**.
- **Indium (In):** It is in the 5th period. The inert pair effect starts to become significant. Both +1 and +3 oxidation states exist, but the +3 state is still generally more stable than the +1 state. However, with large anions like iodide (I^-), the In^{3+} ion has high polarizing power and can oxidize I^- to I_2. InI_3 is less stable than InI. The relationship InI_3 ¿ InI is **incorrect**. In fact, InI is more stable.
- **Thallium (Tl):** It is in the 6th period. The inert pair effect is very strong. The +1 oxidation state is significantly more stable than the +3 oxidation state.
  - Comparing $TlCl_3$ and TlCl: TlCl (+1 state) is more stable than $TlCl_3$ (+3 state). $TlCl_3$ is a strong oxidizing agent and is thermally unstable. The relationship $TlCl_3$ ¿ TlCl is **incorrect**.
  - Comparing TlI_3 and TlI: The +1 state is much more stable. Furthermore, Tl^{3+} is a strong oxidizing agent, and I^- is a good reducing agent. Tl^{3+} would oxidize I^- to I_2. In fact, the compound written as TlI_3 does not contain the Tl^{3+} ion; it exists as $Tl^+(I_3{}^-)$, an ionic compound of thallium(I) and the triiodide ion. Therefore, TlI,

containing the stable Tl^+ ion, is much more stable than a hypothetical TlI_3 containing Tl^{3+}. The relationship TlI ¿ TlI_3 is **correct**.

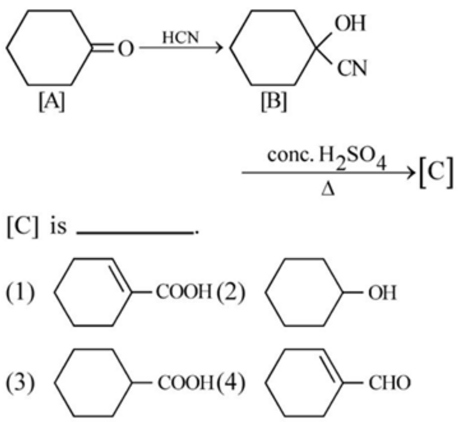
Step 4: Final Answer:
Due to the inert pair effect, the +1 oxidation state is most stable for thallium. Therefore, TlI is more stable than TlI_3.

Quick Tip

Remember the trend for the inert pair effect: stability of the lower oxidation state increases down the group in the p-block. For Group 13, the stability order of the +1 state is Al^+ ¡ Ga^+ ¡ In^+ ¡ Tl^+. For Tl, the +1 state is the most stable.

---

82. Complete the following reaction :



Correct Answer: (A) cyclohex-1-ene-1-carboxylic acid

Solution:

Step 1: Understanding the Question:
The question asks for the final product [C] of a two-step reaction sequence starting from cyclohexanone [A].

Step 2: Key Formula or Approach:
The reaction sequence involves two key transformations: 1. Cyanohydrin formation: A ketone reacts with HCN to form a cyanohydrin. 2. Acid hydrolysis and dehydration: The cyanohydrin is treated with concentrated acid (H_2SO_4) and heat. The nitrile group (-CN) hydrolyzes to a carboxylic acid (-COOH), and the tertiary alcohol group (-OH) undergoes dehydration to form an alkene.

Step 3: Detailed Explanation:
Step I: Formation of Cyanohydrin [B]

The carbonyl group of cyclohexanone [A] is attacked by the nucleophilic cyanide ion (from HCN) to form cyclohexanone cyanohydrin [B].

$$[\text{A: Cyclohexanone}] \xrightarrow{\text{HCN}} [\text{B: 1-hydroxycyclohexanecarbonitrile}]$$

The structure of [B] has both a hydroxyl (-OH) group and a nitrile (-CN) group attached to the same carbon atom (C1) of the ring.

Step II: Formation of Product [C]
Product [B] is heated with concentrated sulfuric acid. Two reactions occur simultaneously:

- **Hydrolysis of Nitrile:** The nitrile group (-C≡N) is completely hydrolyzed by the strong acid to a carboxylic acid group (-COOH).

- **Dehydration of Alcohol:** The hydroxyl group (-OH) is on a tertiary carbon, making it susceptible to dehydration (elimination of a water molecule) in the presence of a strong acid like conc. H_2SO_4 and heat. A double bond is formed between C1 and an adjacent carbon (C2 or C6) of the ring.

The combined result is the formation of cyclohex-1-ene-1-carboxylic acid.

$$[\text{B}] \xrightarrow{\text{conc. } H_2SO_4, \Delta} [\text{C: Cyclohex-1-ene-1-carboxylic acid}] + NH_4HSO_4 + H_2O$$

Step 4: Final Answer:
The final product [C] is cyclohex-1-ene-1-carboxylic acid, which corresponds to the structure shown in option (1).

Quick Tip

When faced with a confusing or seemingly incorrect diagram in a multiple-choice question, try to identify the types of reactions indicated by the reagents (e.g., HCN addition, acid hydrolysis). Then, see if any of the options represent a logical final product for such a reaction type, even if the specific intermediates shown don't make sense.

---

83. The conductivity of centimolar solution of KCl at 25°C is 0.0210 ohm^{-1} cm^{-1} and the resistance of the cell containing the solution at 25°C is 60 ohm. The value of cell constant is :

(A) 1.26 cm^{-1}
(B) 3.34 cm^{-1}
(C) 1.34 cm^{-1}
(D) 3.28 cm^{-1}

Correct Answer: (A) 1.26 cm^{-1}

Solution:

Step 1: Understanding the Question:
The question provides the conductivity (κ) and resistance (R) of a KCl solution in a conductivity cell and asks for the value of the cell constant (G^*).

Step 2: Key Formula or Approach:
The relationship between conductivity (κ), resistance (R), and the cell constant (G^*) is given by the formula:

$$\kappa = \frac{1}{R} \times G^*$$

The cell constant is defined as the ratio of the distance between the electrodes (l) to their cross-sectional area (A), i.e., $G^* = \frac{l}{A}$.

Step 3: Detailed Explanation:
We are given the following values:

- Conductivity, $\kappa = 0.0210\,\Omega^{-1}\text{cm}^{-1}$.
- Resistance, $R = 60\,\Omega$.

We need to find the cell constant, G^*. Rearranging the formula:

$$G^* = \kappa \times R$$

Substitute the given values into the equation:

$$G^* = (0.0210\,\Omega^{-1}\text{cm}^{-1}) \times (60\,\Omega)$$

$$G^* = 0.0210 \times 60 \text{ cm}^{-1}$$

$$G^* = 1.26 \text{ cm}^{-1}$$

The information about the concentration ("centimolar solution") is not required for this specific calculation.

Step 4: Final Answer:
The value of the cell constant is 1.26 cm^{-1}.

Quick Tip

Remember the fundamental equation for conductivity measurements: Conductivity = Conductance $\times$ Cell Constant. Since conductance is the reciprocal of resistance ($G = 1/R$), this becomes $\kappa = (1/R) \times G^*$. This formula is key to solving most problems involving conductivity cells.

---

84. The element expected to form largest ion to achieve the nearest noble gas configuration is :

(A) N
(B) Na

(C) O
(D) F

Correct Answer: (A) N

Solution:

Step 1: Understanding the Question:
The question asks which of the given elements will form the largest ion when it achieves a noble gas configuration.

Step 2: Key Formula or Approach:
1. Determine the stable ion each element forms to achieve a noble gas configuration. 2. Compare the sizes of these ions. 3. For isoelectronic species (ions with the same number of electrons), the ionic radius decreases as the nuclear charge (number of protons) increases. This is because the electrons are pulled more strongly towards the nucleus. 4. For anions, the radius increases with increasing negative charge, as the added electrons increase electron-electron repulsion and are held less tightly by the nucleus.

Step 3: Detailed Explanation:
Let's find the ion formed by each element and its electronic configuration:

- **N (Nitrogen):** Atomic number (Z) = 7. It is in Group 15. To achieve the nearest noble gas configuration (of Neon, 10 electrons), it will gain 3 electrons to form the nitride ion, $\mathbf{N^{3-}}$. It has 7 protons and 10 electrons.
- **Na (Sodium):** Atomic number (Z) = 11. It is in Group 1. To achieve the nearest noble gas configuration (of Neon, 10 electrons), it will lose 1 electron to form the sodium ion, $\mathbf{Na^{+}}$. It has 11 protons and 10 electrons.
- **O (Oxygen):** Atomic number (Z) = 8. It is in Group 16. To achieve the nearest noble gas configuration (of Neon, 10 electrons), it will gain 2 electrons to form the oxide ion, $\mathbf{O^{2-}}$. It has 8 protons and 10 electrons.
- **F (Fluorine):** Atomic number (Z) = 9. It is in Group 17. To achieve the nearest noble gas configuration (of Neon, 10 electrons), it will gain 1 electron to form the fluoride ion, $\mathbf{F^{-}}$. It has 9 protons and 10 electrons.

All four ions formed (N^{3-}, O^{2-}, F^{-}, Na^{+}) are isoelectronic, as they all have 10 electrons.
To compare their sizes, we look at the number of protons (nuclear charge):

- N^{3-}: 7 protons
- O^{2-}: 8 protons
- F^{-}: 9 protons
- Na^{+}: 11 protons

For isoelectronic species, the ion with the fewest protons will have the weakest pull on its electrons, resulting in the largest ionic radius. The N^{3-} ion has only 7 protons holding 10 electrons,

so it will be the largest. The Na^+ ion has 11 protons holding 10 electrons, so it will be the smallest.
The order of ionic radii is: N^{3-} ¿ O^{2-} ¿ F^- ¿ Na^+.

Step 4: Final Answer:
Nitrogen (N) will form the largest ion (N^{3-}) to achieve the nearest noble gas configuration.

Quick Tip

For isoelectronic ions, remember the simple rule: **more protons = smaller ion**. The greater the nuclear charge, the stronger the attraction for the same number of electrons, pulling the electron cloud closer.

---

85. Intermolecular forces are forces of attraction and repulsion between interacting particles that will include :
A. dipole - dipole forces.
B. dipole - induced dipole forces.
C. hydrogen bonding.
D. covalent bonding.
E. dispersion forces.
Choose the most appropriate answer from the options given below :

(A) A, B, C, E are correct.
(B) A, C, D, E are correct.
(C) B, C, D, E are correct.
(D) A, B, C, D are correct.

Correct Answer: (A) A, B, C, E are correct.

Solution:

Step 1: Understanding the Question:
The question asks to identify which forces from the given list are classified as intermolecular forces.

Step 3: Detailed Explanation:
Let's define each force type and determine if it is intermolecular or intramolecular.

- **Intermolecular forces** (IMFs) are forces that exist *between* molecules. They are generally weaker than intramolecular forces.
- **Intramolecular forces** are forces that exist *within* a molecule, holding the atoms together (i.e., chemical bonds).

Now let's classify the given forces:

- **A. dipole - dipole forces:** These are attractive forces between the positive end of one polar molecule and the negative end of another polar molecule. This is an **intermolecular force**.
- **B. dipole - induced dipole forces:** These occur when a polar molecule induces a temporary dipole in a nonpolar molecule, leading to an attraction. This is an **intermolecular force**.
- **C. hydrogen bonding:** This is a special, strong type of dipole-dipole interaction that occurs between a hydrogen atom bonded to a highly electronegative atom (N, O, or F) and another nearby electronegative atom. This is an **intermolecular force**.
- **D. covalent bonding:** This is the force that holds atoms together within a molecule through the sharing of electrons. This is an **intramolecular force**, not an intermolecular one.
- **E. dispersion forces (London forces):** These are weak forces arising from temporary, induced dipoles in atoms or molecules. They are present between all types of molecules. This is an **intermolecular force**.

The forces classified as intermolecular forces are A, B, C, and E. Covalent bonding (D) is an intramolecular force. Therefore, we should choose the option that includes A, B, C, and E, but excludes D.

Step 4: Final Answer:
The correct set of intermolecular forces is A, B, C, and E. This corresponds to option (A).

Quick Tip

A simple way to distinguish is to ask: "Does this force hold a molecule together, or does it attract one molecule to another?" Covalent, ionic, and metallic bonds hold molecules/crystals together (intramolecular). Van der Waals forces (dipole-dipole, London) and hydrogen bonds attract separate molecules to each other (intermolecular).

---

86. The reaction that does NOT take place in a blast furnace between 900 K to 1500 K temperature range during extraction of iron is :

(1) $C + CO_2 \rightarrow 2CO$
(2) $CaO + SiO_2 \rightarrow CaSiO_3$
(3) $Fe_2O_3 + CO \rightarrow 2FeO + CO_2$
(4) $FeO + CO \rightarrow Fe + CO_2$

Correct Answer: (3) $Fe_2O_3 + CO \rightarrow 2FeO + CO_2$

Solution:

Step 1: Understanding the Question:
The question asks to identify which of the given chemical reactions does not occur in the higher temperature zone (900 K to 1500 K) of a blast furnace used for iron extraction.

Step 2: Detailed Explanation:
The blast furnace has different temperature zones where specific reactions occur.
Lower Temperature Zone (500 K - 800 K): This is the upper part of the furnace. Here, the iron ore (mainly Fe_2O_3) is reduced in steps by carbon monoxide (CO).

$$3Fe_2O_3 + CO \rightarrow 2Fe_3O_4 + CO_2$$

$$Fe_3O_4 + CO \rightarrow 3FeO + CO_2$$

$$FeO + CO \rightarrow Fe + CO_2$$

The reaction $Fe_2O_3 + CO \rightarrow 2FeO + CO_2$ is a summary of the initial reduction steps that occur at these lower temperatures.

Higher Temperature Zone (900 K - 1500 K): This is the central and lower part of the furnace.
- **Slag Formation:** Limestone ($CaCO_3$) decomposes to CaO, which then reacts with silica impurities (SiO_2) to form molten slag ($CaSiO_3$). This happens around 1200 K. So, reaction (2) **occurs** in this range.

$$CaO + SiO_2 \rightarrow CaSiO_3$$

- **CO Generation (Boudouard reaction):** Carbon dioxide reacts with hot coke to form carbon monoxide. This reaction is endothermic and favored at high temperatures (> 1000 K). So, reaction (1) **occurs** in this range.

$$C + CO_2 \rightarrow 2CO$$

- **Final Reduction:** The final reduction of iron oxide (FeO) to molten iron occurs here, primarily by CO and also by direct reduction with carbon. So, reaction (4) also **occurs** in this range.

$$FeO + CO \rightarrow Fe + CO_2$$

Step 3: Final Answer:
Based on the analysis of the temperature zones, the reduction of Fe_2O_3 (reaction 3) primarily takes place in the cooler, upper part of the furnace, at temperatures below 900 K. Therefore, it is the reaction that does not take place in the 900 K to 1500 K range.

Quick Tip

To remember the reactions in a blast furnace, think of it as a top-to-bottom process. The ore enters at the top (cooler region) and gets pre-heated and partially reduced. As it moves down, the temperature increases, and the final reduction and melting occur at the bottom (hottest region).

87. Match List - I with List - II :
List - I (Oxoacids of Sulphur)
A. Peroxodisulphuric acid
B. Sulphuric acid
C. Pyrosulphuric acid
D. Sulphurous acid
List - II (Bonds)
I. Two S-OH, Four S=O, One S-O-S
II. Two S-OH, One S=O
III. Two S-OH, Four S=O, One S-O-O-S
IV. Two S-OH, Two S=O
Choose the correct answer from the options given below :

(1) A-I, B-III, C-IV, D-II
(2) A-III, B-IV, C-II, D-I
(3) A-I, B-III, C-II, D-IV
(4) A-III, B-IV, C-I, D-II

Correct Answer: (4) A-III, B-IV, C-I, D-II

Solution:

Step 1: Understanding the Question:
The task is to match each oxoacid of sulphur from List-I with the correct description of its chemical bonds from List-II. This requires knowledge of the structures of these acids.

Step 2: Analyzing the Structure of Each Oxoacid:
A. Peroxodisulphuric acid ($H_2S_2O_8$, Marshall's acid):
The structure contains a peroxide linkage (-O-O-) connecting two SO_3H groups.
Structure: HO-SO_2-O-O-SO_2-OH.
Bonds: Two S-OH bonds, four S=O bonds, and one S-O-O-S chain (which implies a central peroxide bond). This matches description **III**.

B. Sulphuric acid (H_2SO_4):
The structure has a central sulphur atom double-bonded to two oxygen atoms and single-bonded to two hydroxyl (-OH) groups.
Structure: $(HO)_2SO_2$.
Bonds: Two S-OH bonds and two S=O bonds. This matches description **IV**.

C. Pyrosulphuric acid ($H_2S_2O_7$, Oleum):
This acid is formed by joining two sulphuric acid molecules with the removal of one water molecule. It contains an S-O-S linkage.
Structure: HO-SO_2-O-SO_2-OH.
Bonds: Two S-OH bonds, four S=O bonds, and one S-O-S bridge. This matches description **I**.

D. Sulphurous acid (H_2SO_3):
The structure has a central sulphur atom double-bonded to one oxygen atom and single-bonded to two hydroxyl (-OH) groups. Sulphur also has a lone pair of electrons.
Structure: $(HO)_2SO$.
Bonds: Two S-OH bonds and one S=O bond. This matches description **II**.

Step 3: Matching and Final Answer:
Based on the analysis:
- A matches with III.
- B matches with IV.
- C matches with I.
- D matches with II.

The correct combination is A-III, B-IV, C-I, D-II, which corresponds to option (4).

> Quick Tip
>
> Drawing the Lewis structures for oxoacids is the key to solving such matching problems. Remember the common linkages: S-O-S for 'pyro' acids and -O-O- for 'peroxo' or 'peroxy' acids.

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88. Pumice stone is an example of -

(1) solid sol
(2) foam
(3) sol
(4) gel

Correct Answer: (2) foam

Solution:

Step 1: Understanding the Question:
The question asks to classify pumice stone based on its colloidal nature.

Step 2: Detailed Explanation:
Colloidal systems are classified based on the physical state of the dispersed phase and the dispersion medium. Let's analyze the structure of pumice stone and the given options.
Pumice stone: It is a volcanic rock formed when super-heated, highly pressurized rock is violently ejected from a volcano. The porous texture is due to gas bubbles being trapped in the rock as it cooled. Therefore, it consists of a gaseous phase (dispersed phase) trapped within a solid phase (dispersion medium).

Let's define the given colloid types:
- **Solid sol:** Dispersed phase is solid, dispersion medium is solid (e.g., colored gemstones).

- **Foam:** Dispersed phase is gas, dispersion medium is liquid (e.g., whipped cream). A system with gas dispersed in a solid is called a **solid foam**.
- **Sol:** Dispersed phase is solid, dispersion medium is liquid (e.g., paint).
- **Gel:** A liquid is dispersed in a solid medium in such a way that it forms a semi-solid network (e.g., jelly, cheese).

Step 3: Final Answer:
Since pumice stone is a dispersion of gas in a solid, it is an example of a solid foam. The general term "foam" is used in the options to represent this class of colloids. Therefore, option (2) is the correct answer.

Quick Tip

Create a table to remember the types of colloids. For example: Dispersed Phase / Dispersion Medium → Colloid Type (Example). Gas / Solid → Solid foam (Pumice stone, Styrofoam). This makes recalling examples easier.

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89. Given below are two statements :
Statement I: The nutrient deficient water bodies lead to eutrophication.
Statement II: Eutrophication leads to decrease in the level of oxygen in the water bodies.
In the light of the above statements, choose the correct answer from the options given below :

(1) Statement I is correct but Statement II is false.
(2) Statement I is incorrect but Statement II is true.
(3) Both Statement I and Statement II are true.
(4) Both Statement I and Statement II are false.

Correct Answer: (2) Statement I is incorrect but Statement II is true.

Solution:

Step 1: Understanding the Question:
The question presents two statements about the environmental phenomenon of eutrophication and asks to evaluate their correctness.

Step 2: Detailed Explanation:
Analysis of Statement I:
"The nutrient deficient water bodies lead to eutrophication."
Eutrophication is the process of nutrient enrichment of a water body, typically with nitrates and phosphates. This excess of nutrients promotes excessive growth of algae and other aquatic plants (algal bloom). Therefore, eutrophication is caused by nutrient-rich, not nutrient-deficient, con-

ditions. Thus, Statement I is incorrect.

Analysis of Statement II:
"Eutrophication leads to decrease in the level of oxygen in the water bodies."
The algal blooms caused by eutrophication eventually die off. The dead organic matter is then decomposed by aerobic bacteria. This decomposition process consumes large amounts of dissolved oxygen from the water. The depletion of oxygen can lead to hypoxic (low oxygen) or anoxic (no oxygen) conditions, which is harmful to fish and other aquatic life. Thus, Statement II is correct.

Step 3: Final Answer:
Based on the analysis, Statement I is incorrect and Statement II is true. This corresponds to option (2).

Quick Tip

Remember that "Eutrophication" means "well-nourished". This term directly points to nutrient enrichment, not deficiency. The subsequent oxygen depletion is a key consequence of this process.

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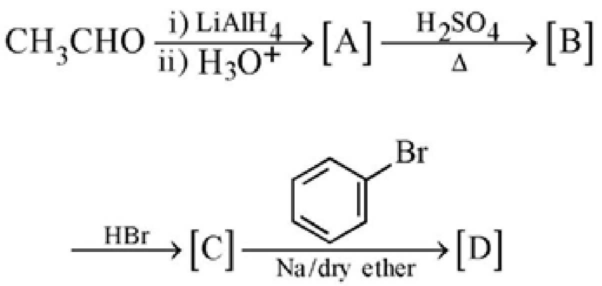
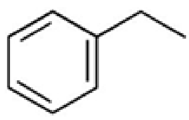
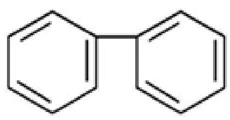
90. Identify the final product [D] obtained in the following sequence of reactions.



(1) $HC\equiv C^{\ominus}\ Na^{+}$

(2) 

(3) 

(4) C_4H_{10}

(A) $HC\equiv C^{-}Na^{+}$
(B) Ethylbenzene

(C) Biphenyl
(D) C_4H_{10}

Correct Answer: (2) Ethylbenzene

Solution:

Step 1: Understanding the Question:
The question presents a multi-step reaction sequence and asks for the structure of the final product [D].

Step 2: Key Formula or Approach:
We need to identify the product of each step in the sequence. The key reactions are reduction of an aldehyde, dehydration of an alcohol, addition of HBr to an alkene, and a Wurtz-Fittig reaction.

Step 3: Detailed Explanation:

- **Step 1:** $\mathbf{CH_3CHO \xrightarrow{i)\ LiAlH_4\ ii)\ H_3O^+} [A]}$
  Lithium aluminium hydride ($LiAlH_4$) is a strong reducing agent that reduces the aldehyde ethanal (CH_3CHO) to a primary alcohol.
  $$[A] \text{ is } CH_3CH_2OH \text{ (Ethanol)}$$
- **Step 2:** $\mathbf{[A] \xrightarrow{H_2SO_4, \Delta} [B]}$
  Ethanol is dehydrated by concentrated sulfuric acid upon heating to form an alkene.
  $$[B] \text{ is } CH_2 = CH_2 \text{ (Ethene)}$$
- **Step 3:** $\mathbf{[B] \xrightarrow{HBr} [C]}$
  Ethene undergoes electrophilic addition with hydrogen bromide.
  $$[C] \text{ is } CH_3CH_2Br \text{ (Bromoethane)}$$
- **Step 4:** $\mathbf{[C] + \text{Bromobenzene} \xrightarrow{Na/dry\ ether} [D]}$
  This is a Wurtz-Fittig reaction. Bromoethane (an alkyl halide) reacts with bromobenzene (an aryl halide) and sodium metal in dry ether to form an alkylbenzene. The ethyl group from bromoethane attaches to the phenyl ring from bromobenzene.
  $$CH_3CH_2Br + C_6H_5Br + 2Na \xrightarrow{\text{dry ether}} C_6H_5\text{-}CH_2CH_3 + 2NaBr$$
  $$[D] \text{ is Ethylbenzene}$$

Step 4: Final Answer:
The final product [D] is ethylbenzene, which is represented by the structure in option (2).

Quick Tip

Pay close attention to the layout of reaction schemes. The reaction of an alkyl halide and an aryl halide with sodium in ether is a specific named reaction called the Wurtz-Fittig reaction, which is used to synthesize alkylbenzenes.

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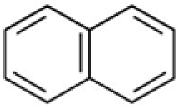
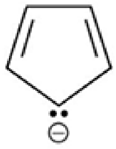
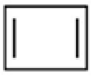
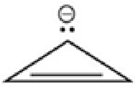
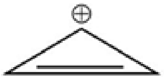
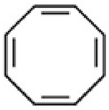
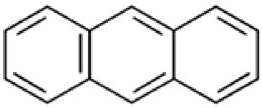
91. Consider the following compounds/species:

i.  ii. 

iii.  iv. 

v.  vi. 

vii. 

The number of compounds/species which obey Huckel's rule is ______.

(1) 2
(2) 5
(3) 4
(4) 6

Correct Answer: (3) 4

Solution:

Step 1: Understanding the Question:
The question asks us to count how many of the given seven species are aromatic based on Hückel's rule.

Step 2: Key Formula or Approach:
Hückel's rule states that for a species to be aromatic, it must satisfy four conditions: 1. It must be cyclic. 2. It must be planar. 3. It must be completely conjugated (every atom in the ring must have a p-orbital). 4. It must contain $(4n + 2)$ π electrons, where n is a non-negative integer (0, 1, 2, ...).

Step 3: Detailed Explanation:
Let's analyze each species:

- **i. Naphthalene:** It is cyclic, planar, and fully conjugated. It has 10 π electrons. For 4n + 2 = 10, 4n = 8, so n = 2. It obeys Hückel's rule. **(Aromatic)**

- **ii. Cyclopentadienyl anion:** It is cyclic, planar, and fully conjugated. It has 6 π electrons (4 from double bonds, 2 from the lone pair/negative charge). For 4n + 2 = 6, 4n = 4, so n = 1. It obeys Hückel's rule. **(Aromatic)**

- **iii. Cyclopropenyl cation:** It is a three-membered ring with a positive charge. It is cyclic, planar, and fully conjugated. It has 2 π electrons. For 4n + 2 = 2, 4n = 0, so n = 0. It obeys Hückel's rule. **(Aromatic)**

- **iv. Bicyclo[1.1.0]butane:** This is a bicyclic, non-planar molecule. It is not aromatic. **(Non-aromatic)**

- **v. Cyclopropenyl cation:** This appears to be the same as species iii. Assuming it's a distinct species intended, like cyclobutadiene, which has 4π electrons (anti-aromatic), or some other non-aromatic species, it doesn't add to the count. Let's assume it is just a repeated structure.

- **vi. Cyclooctatetraene (COT):** It is cyclic and has 8 π electrons (a 4n system, where n=2). To avoid the instability of being anti-aromatic, it adopts a non-planar, tub-like shape. Since it's not planar, it is not aromatic. **(Non-aromatic)**

- **vii. Anthracene:** It is cyclic, planar, and fully conjugated. It has 14 π electrons. For 4n + 2 = 14, 4n = 12, so n = 3. It obeys Hückel's rule. **(Aromatic)**

Step 4: Final Answer:
The species that are aromatic are i, ii, iii, and vii. Counting these, we find there are 4 aromatic species. This corresponds to option (3).

> **Quick Tip**
>
> When applying Hückel's rule, check all four conditions: cyclic, planar, conjugated, and the $(4n + 2)$ π electron count. Be aware that polycyclic systems like naphthalene and anthracene are also considered aromatic and are generally treated as obeying the rule in this context. If your count doesn't match the options, re-check for potential ambiguities or common exam question errors.

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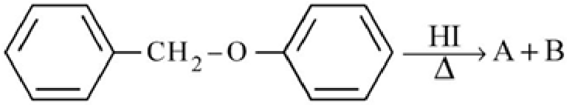
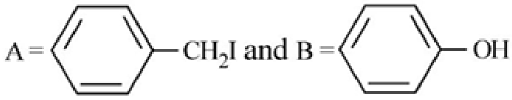
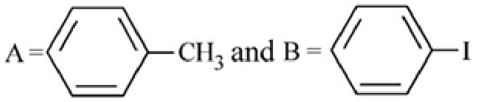
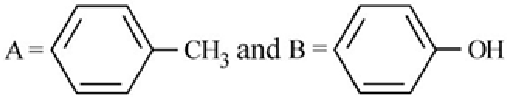
92. Consider the following reaction :

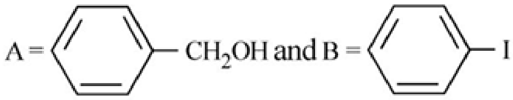


Identify products A and B.

(1) 

(2) 

(3) 

(4) 

Correct Answer: (1) Option A

Solution:

Step 1: Understanding the Question:
The question shows the reaction of benzyl phenyl ether with hydrogen iodide (HI) and asks to identify the two products, A and B. This is a classic example of ether cleavage.

Step 2: Key Formula or Approach:
The cleavage of ethers by strong acids like HI involves two main steps: 1. Protonation of the ether oxygen to make it a better leaving group. 2. Nucleophilic attack by the iodide ion (I^-) on one of the carbon atoms attached to the oxygen, breaking the C-O bond. The attack follows either an S_N1 or S_N2 mechanism depending on the structure of the ether.

Step 3: Detailed Explanation:
The starting material is benzyl phenyl ether: C_6H_5-O-CH_2-C_6H_5. The ether oxygen is bonded to a phenyl group (an sp^2-hybridized carbon) and a benzyl group (an sp^3-hybridized carbon). There are two possible C-O bonds to cleave:
1. **Aryl-Oxygen bond (C_6H_5–O):** This bond is very strong and difficult to break. The carbon atom of the benzene ring is sp^2 hybridized and has partial double-bond character due to resonance with the oxygen lone pairs. S_N2 attack is not possible, and formation of a phenyl cation for an S_N1 reaction is extremely unfavorable.
2. **Benzyl-Oxygen bond (-$CH_2C_6H_5$–O):** This is a standard alkyl-oxygen single bond. The benzylic carbon is a good site for nucleophilic substitution. The iodide ion (I^-) will attack the less sterically hindered and more electrophilic benzylic carbon.
The reaction proceeds as follows: - The oxygen is protonated by HI. - The I^- ion attacks the

benzylic carbon (-CH_2-), breaking the C-O bond. - The products formed are benzyl iodide ($C_6H_5CH_2I$) and phenol (C_6H_5OH).

Step 4: Final Answer:
The products are benzyl iodide and phenol. Comparing this with the options, option (1) correctly identifies A as benzyl iodide and B as phenol.

Quick Tip

A key rule for ether cleavage with HX: an aryl-oxygen bond in an aryl alkyl ether is never cleaved. The halogen always attaches to the alkyl group, and a phenol is formed.

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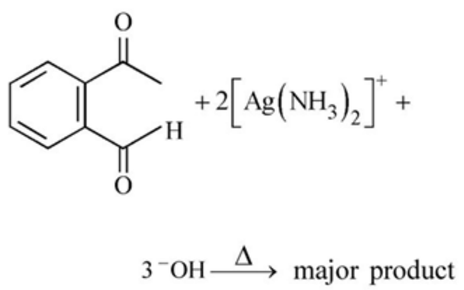
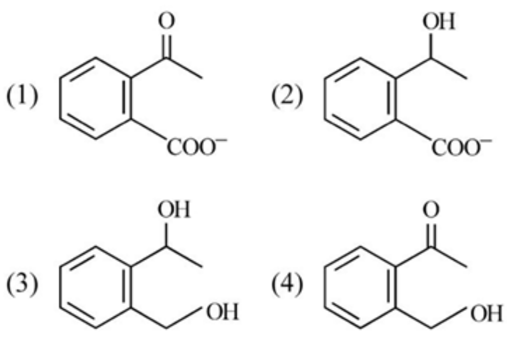
93. Identify the major product obtained in the following reaction :





Correct Answer: (1) The structure corresponding to 2-formylbenzoate.

Solution:

Step 1: Understanding the Question:
The reaction shows o-phthalaldehyde reacting with Tollens' reagent ($[Ag(NH_3)_2]^+$) in a basic medium (OH^-). We need to identify the major product.

Step 2: Analyzing the Reactants and Reagents:
Reactant: o-phthalaldehyde, which has two aldehyde (-CHO) groups on a benzene ring at adjacent positions.

Reagent: Tollens' reagent ($[Ag(NH_3)_2]^+$). This is a mild oxidizing agent used to test for aldehydes. It oxidizes an aldehyde to a carboxylate anion (R-CHO $\rightarrow$ R-COO^-).
Stoichiometry: The reaction specifies '+ $2[Ag(NH_3)_2]^+$'. The balanced reaction for the oxidation of one aldehyde group is:

$$\text{R-CHO} + 2[Ag(NH_3)_2]^+ + 3OH^- \rightarrow \text{R-COO}^- + 2Ag(s) + 4NH_3 + 2H_2O$$

This shows that 2 moles of Tollens' reagent are required to oxidize 1 mole of an aldehyde group.

Step 3: Applying Stoichiometry to the Reaction:
The starting material, o-phthalaldehyde, has two aldehyde groups. To oxidize both groups, we would need $2 \times 2 = 4$ moles of $[Ag(NH_3)_2]^+$.
However, the question explicitly provides only 2 moles of $[Ag(NH_3)_2]^+$. This means the Tollens' reagent is the limiting reagent, and it has just enough quantity to oxidize only one of the two aldehyde groups.
Therefore, one -CHO group will be oxidized to a -COO^- group, while the other -CHO group will remain unchanged.

$$C_6H_4(CHO)_2 + 2[Ag(NH_3)_2]^+ + 3OH^- \rightarrow C_6H_4(CHO)(COO^-) + 2Ag(s) + 4NH_3 + 2H_2O$$

Step 4: Final Answer:
The major product is the 2-formylbenzoate anion. This structure is shown in option (1). Note that other reactions like the Cannizzaro reaction are possible in basic medium, but the presence and specific stoichiometry of the oxidizing agent (Tollens' reagent) dictate the course of this reaction.

Quick Tip

In organic chemistry reactions, always pay close attention to the stoichiometry of the reagents provided. It can act as a crucial hint to determine the product, especially when multiple reactive sites are present in the substrate. Here, the limited amount of oxidant leads to partial oxidation.

---

94. Which complex compound is most stable?

(1) $[CoCl_2(en)_2]NO_3$
(2) $[Co(NH_3)_6]_2(SO_4)_3$
(3) $[Co(NH_3)_4(H_2O)Br](NO_3)_2$
(4) $[Co(NH_3)_3(NO_3)_3]$

Correct Answer: (1) $[CoCl_2(en)_2]NO_3$

Solution:

Step 1: Understanding the Question:
The question asks to identify the most stable coordination compound among the given options.

Step 2: Key Formula or Approach:
The stability of a coordination compound is significantly influenced by the **chelate effect**. The chelate effect refers to the enhanced stability of complexes containing chelating ligands (ligands that can bind to the central metal ion through more than one donor atom, forming a ring). These complexes, called chelates, are more stable than similar complexes with monodentate ligands.

Step 3: Detailed Explanation:
Let's analyze the ligands in each complex:
- **(1) $[CoCl_2(en)_2]NO_3$:** The ligands are Cl^- (monodentate) and 'en' which stands for ethylenediamine (NH_2-CH_2-CH_2-NH_2). Ethylenediamine is a bidentate ligand, meaning it binds to the cobalt ion at two points, forming a stable five-membered ring. This is a chelate.
- **(2) $[Co(NH_3)_6]_2(SO_4)_3$:** The ligand is NH_3 (ammonia), which is a monodentate ligand. No chelate rings are formed.
- **(3) $[Co(NH_3)_4(H_2O)Br](NO_3)_2$:** The ligands are NH_3, H_2O, and Br^-, all of which are monodentate. No chelate rings are formed.
- **(4) $[Co(NH_3)_3(NO_3)_3$:]** The ligands are NH_3 and NO_3^- (acting as a monodentate ligand here). No chelate rings are formed.

Step 4: Final Answer:
Since the complex in option (1) is the only one containing a chelating ligand (ethylenediamine), it will be the most stable due to the chelate effect. The formation of these rings increases the thermodynamic stability of the complex.

Quick Tip

When comparing the stability of coordination compounds, always look for the presence of polydentate (chelating) ligands like ethylenediamine (en), oxalate (ox), EDTA, etc. Complexes with these ligands are almost always more stable than those with only monodentate ligands.

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95. The equilibrium concentrations of the species in the reaction $A + B \rightleftharpoons C + D$ are 2, 3, 10 and 6 mol L^{-1}, respectively at 300 K. ΔG° for the reaction is (R = 2 cal / mol K)

(1) -1381.80 cal
(2) -13.73 cal
(3) 1372.60 cal
(4) -137.26 cal

Correct Answer: (1) -1381.80 cal

Solution:

Step 1: Understanding the Question:
We are given the equilibrium concentrations for a reaction and asked to calculate the standard Gibbs free energy change (ΔG°).

Step 2: Key Formula or Approach:
The relationship between the standard Gibbs free energy change and the equilibrium constant (K) is given by:

$$\Delta G^\circ = -RT\ln(K)$$

First, we need to calculate the equilibrium constant (K_c) from the given concentrations. For the reaction A + B $\rightleftharpoons$ C + D, the expression for K_c is:

$$K_c = \frac{[\mathrm{C}][\mathrm{D}]}{[\mathrm{A}][\mathrm{B}]}$$

Step 3: Detailed Explanation:
Part A: Calculate the equilibrium constant (K_c)
Given equilibrium concentrations:
A
= 2 mol L^{-1}
B
= 3 mol L^{-1}
C
= 10 mol L^{-1}
D
= 6 mol L^{-1}
Substituting these values into the K_c expression:

$$K_c = \frac{(10)(6)}{(2)(3)} = \frac{60}{6} = 10$$

Part B: Calculate ΔG°
Given values:
R = 2 cal / mol K
T = 300 K
K_c = 10
Using the formula $\Delta G^\circ = -RT\ln(K)$:

$$\Delta G^\circ = -(2 \text{ cal/mol K}) \times (300 \text{ K}) \times \ln(10)$$
$$\Delta G^\circ = -600 \times \ln(10) \text{ cal/mol}$$

We know that $\ln(10) \approx 2.303$.

$$\Delta G^\circ = -600 \times 2.303 \text{ cal/mol}$$
$$\Delta G^\circ = -1381.8 \text{ cal/mol}$$

Step 4: Final Answer:
The standard Gibbs free energy change for the reaction is -1381.80 cal. This matches option (1).

Quick Tip

Remember the sign convention for ΔG°. If K ¿ 1, the reaction favors products, $\ln(K)$ is positive, and ΔG° is negative (spontaneous under standard conditions). If K ¡ 1, $\ln(K)$ is negative, and ΔG° is positive.

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96. What fraction of one edge centred octahedral void lies in one unit cell of fcc?

(1) $\frac{1}{4}$
(2) $\frac{1}{12}$
(3) $\frac{1}{2}$
(4) $\frac{1}{3}$

Correct Answer: (1) $\frac{1}{4}$

Solution:

Step 1: Understanding the Question:
The question asks for the contribution of a single octahedral void located at the center of an edge to one face-centered cubic (fcc) unit cell.

Step 2: Location of Octahedral Voids in FCC:
In an fcc lattice, octahedral voids are located at two types of positions:
1. At the body center of the cube.
2. At the center of each of the 12 edges of the cube.

Step 3: Contribution of an Edge-Centered Void:
An atom or void located at the center of an edge of a cubic unit cell is shared by the four unit cells that meet at that edge. Imagine a cube; one edge is also the edge for three other cubes (one adjacent, one above, and one adjacent to the one above).
Therefore, any particle or void at an edge center contributes only $\frac{1}{4}$ of its volume to a single unit cell.

Step 4: Final Answer:
The fraction of one edge-centered octahedral void that lies within one unit cell is $\frac{1}{4}$.

Quick Tip

To solve problems on solid state, visualize the unit cell and remember the contribution of particles at different positions: Corner = 1/8, Face Center = 1/2, Body Center = 1, Edge Center = 1/4. This applies to both atoms and voids.

---

97. Which amongst the following options is the correct relation between change in enthalpy and change in internal energy?

(1) $\Delta H - \Delta U = -\Delta n_g RT$
(2) $\Delta H + \Delta U = \Delta n_g R$
(3) $\Delta H = \Delta U - \Delta n_g RT$
(4) $\Delta H = \Delta U + \Delta n_g RT$

Correct Answer: (4) $\Delta H = \Delta U + \Delta n_g RT$

Solution:

Step 1: Understanding the Question:
The question asks for the mathematical relationship between the change in enthalpy (ΔH) and the change in internal energy (ΔU) for a chemical reaction.

Step 2: Key Formula or Approach:
The definition of enthalpy (H) is given by the equation:

$$H = U + PV$$

where U is the internal energy, P is the pressure, and V is the volume.
For a change in the system, this relationship can be written as:

$$\Delta H = \Delta U + \Delta(PV)$$

Step 3: Detailed Explanation:
For a chemical reaction involving gases, we can assume ideal gas behavior. The ideal gas equation is $PV = nRT$.
Substituting this into the enthalpy change equation:

$$\Delta H = \Delta U + \Delta(nRT)$$

If the reaction occurs at a constant temperature (T), R and T can be taken out of the delta operator:

$$\Delta H = \Delta U + RT\Delta n$$

Here, Δn represents the change in the number of moles of gaseous components in the reaction. It is denoted as Δn_g.

$$\Delta n_g = (\text{moles of gaseous products}) - (\text{moles of gaseous reactants})$$

Thus, the final relationship is:

$$\Delta H = \Delta U + \Delta n_g RT$$

Step 4: Final Answer:
Comparing this derived equation with the given options, we find that option (4) is the correct

relation.

Quick Tip

When using the formula $\Delta H = \Delta U + \Delta n_g RT$, remember that Δn_g only includes the change in the number of moles of **gaseous** substances. Moles of solids and liquids are not considered in this term.

---

98. On balancing the given redox reaction,
a $Cr_2O_7^{2-}$ (aq) + b SO_3^{2-} (aq) + c H^+ (aq) $\rightarrow$ 2a Cr^{3+} (aq) + b SO_4^{2-} (aq) + $\frac{c}{2}$ H_2O (l)
the coefficients a, b and c are found to be, respectively -

(1) 1, 8, 3
(2) 8, 1, 3
(3) 1, 3, 8
(4) 3, 8, 1

Correct Answer: (3) 1, 3, 8

Solution:

Step 1: Understanding the Question:
The task is to balance the given redox reaction in an acidic medium and find the values of the stoichiometric coefficients a, b, and c.

Step 2: Key Formula or Approach:
We will use the ion-electron method (half-reaction method) to balance the equation.
1. Separate the overall reaction into oxidation and reduction half-reactions.
2. Balance atoms other than O and H.
3. Balance O atoms by adding H_2O.
4. Balance H atoms by adding H^+.
5. Balance the charge by adding electrons (e^-).
6. Multiply the half-reactions by integers to make the number of electrons equal in both.
7. Add the balanced half-reactions and cancel out common species.

Step 3: Detailed Explanation:
The unbalanced reaction is: $Cr_2O_7^{2-} + SO_3^{2-} \rightarrow Cr^{3+} + SO_4^{2-}$

Reduction Half-Reaction:
$Cr_2O_7^{2-} \rightarrow Cr^{3+}$
- Balance Cr: $Cr_2O_7^{2-} \rightarrow 2Cr^{3+}$
- Balance O: $Cr_2O_7^{2-} \rightarrow 2Cr^{3+} + 7H_2O$
- Balance H: $Cr_2O_7^{2-} + 14H^+ \rightarrow 2Cr^{3+} + 7H_2O$

- Balance charge: The left side has a charge of (-2) + (+14) = +12. The right side has 2 × (+3) = +6. Add $6e^-$ to the left side.
$\mathbf{Cr_2O_7^{2-} + 14H^+ + 6e^- \rightarrow 2Cr^{3+} + 7H_2O}$

Oxidation Half-Reaction:
$SO_3^{2-} \rightarrow SO_4^{2-}$
- S is already balanced.
- Balance O: $SO_3^{2-} + H_2O \rightarrow SO_4^{2-}$
- Balance H: $SO_3^{2-} + H_2O \rightarrow SO_4^{2-} + 2H^+$
- Balance charge: The left side has a charge of -2. The right side has (-2) + (+2) = 0. Add $2e^-$ to the right side.
$\mathbf{SO_3^{2-} + H_2O \rightarrow SO_4^{2-} + 2H^+ + 2e^-}$

Combine the Half-Reactions:
To equalize the electrons, multiply the oxidation half-reaction by 3.
Reduction: $Cr_2O_7^{2-} + 14H^+ + 6e^- \rightarrow 2Cr^{3+} + 7H_2O$
Oxidation (× 3): $3SO_3^{2-} + 3H_2O \rightarrow 3SO_4^{2-} + 6H^+ + 6e^-$
Adding them together:
$Cr_2O_7^{2-} + 14H^+ + 3SO_3^{2-} + 3H_2O \rightarrow 2Cr^{3+} + 7H_2O + 3SO_4^{2-} + 6H^+$
Cancel common species (H^+, H_2O, e^-):
$\mathbf{1\ Cr_2O_7^{2-} + 3\ SO_3^{2-} + 8\ H^+ \rightarrow 2\ Cr^{3+} + 3\ SO_4^{2-} + 4\ H_2O}$

Step 4: Final Answer:
Comparing the balanced equation with the given format:
a $Cr_2O_7^{2-}$ + b SO_3^{2-} + c $H^+ \rightarrow$ ...
We find that **a = 1**, **b = 3**, and **c = 8**.
This corresponds to option (3).

Quick Tip

After balancing a redox reaction, always perform a final check of both atom balance and charge balance. In the final equation: Left charge = (1 × -2) + (3 × -2) + (8 × +1) = -2 - 6 + 8 = 0. Right charge = (2 × +3) + (3 × -2) = +6 - 6 = 0. The charges are balanced.

---

99. Which of the following statements are INCORRECT?
A. All the transition metals except scandium form MO oxides which are ionic.
B. The highest oxidation number corresponding to the group number in transition metal oxides is attained in Sc_2O_3 to Mn_2O_7.
C. Basic character increases from V_2O_3 to V_2O_4 to V_2O_5.
D. V_2O_4 dissolves in acids to give VO_4^{3-} salts.
E. CrO is basic but Cr_2O_3 is amphoteric.
Choose the correct answer from the options given below :

(1) C and D only
(2) B and C only
(3) A and E only
(4) B and D only

Correct Answer: (1) C and D only

Solution:

Step 1: Understanding the Question:
The question asks to identify the incorrect statements among the given options related to the properties of transition metal oxides. We need to analyze each statement individually.

Step 2: Detailed Explanation of Each Statement:
Statement A: All the transition metals except scandium form MO oxides which are ionic.
Scandium (Sc) primarily exists in the +3 oxidation state, forming Sc_2O_3, not ScO. Other first-row transition metals (Ti, V, Cr, Mn, Fe, Co, Ni, Cu) do form oxides of the type MO (e.g., TiO, VO, FeO, CuO). In these oxides, the metal is in a lower oxidation state (+2), and these oxides are predominantly ionic and basic in nature. So, statement A is correct.

Statement B: The highest oxidation number corresponding to the group number in transition metal oxides is attained in Sc_2O_3 to Mn_2O_7.
Let's check the group number and highest oxidation state for elements from Sc to Mn:
- Scandium (Group 3): Highest oxidation state is +3 (in Sc_2O_3). Correct.
- Titanium (Group 4): Highest oxidation state is +4 (in TiO_2). Correct.
- Vanadium (Group 5): Highest oxidation state is +5 (in V_2O_5). Correct.
- Chromium (Group 6): Highest oxidation state is +6 (in CrO_3). Correct.
- Manganese (Group 7): Highest oxidation state is +7 (in Mn_2O_7). Correct.
This trend holds true up to manganese. After manganese, the highest oxidation state is generally lower than the group number (e.g., Fe in Group 8 has a maximum oxidation state of +6, not +8). The statement correctly describes the range from Sc to Mn. So, statement B is correct.

Statement C: Basic character increases from V_2O_3 to V_2O_4 to V_2O_5.
The acidic/basic character of metal oxides depends on the oxidation state of the metal. As the oxidation state increases, the covalent character of the M-O bond increases, and the oxide becomes more acidic.
- V_2O_3 (V is +3): Basic
- V_2O_4 (V is +4): Amphoteric
- V_2O_5 (V is +5): Acidic
Therefore, the basic character *decreases* as we go from V_2O_3 to V_2O_5. The statement claims it increases. So, statement C is incorrect.

Statement D: V_2O_4 dissolves in acids to give VO_4^{3-} salts.
V_2O_4 is an oxide where Vanadium is in the +4 oxidation state. When it dissolves in acids, it forms vanadyl salts containing the vanadyl ion, $[VO]^{2+}$, where V is still in the +4 oxidation state. The ion mentioned, VO_4^{3-} (orthovanadate), has Vanadium in the +5 oxidation state.

Dissolving in a non-oxidizing acid does not change the oxidation state. Therefore, the product formed is incorrect. So, statement D is incorrect.

Statement E: CrO is basic but Cr_2O_3 is amphoteric.
Similar to vanadium, the character of chromium oxides depends on the oxidation state.
- CrO (Cr is +2): Basic
- Cr_2O_3 (Cr is +3): Amphoteric
- CrO_3 (Cr is +6): Acidic
The statement is consistent with this trend. So, statement E is correct.

Step 3: Final Answer:
The incorrect statements are C and D. Therefore, the correct option is (1).

Quick Tip

Remember the general trend for metal oxides: as the oxidation state of the metal increases, the ionic character decreases, and the acidic character increases. Lower oxidation state oxides are basic, intermediate ones are amphoteric, and higher oxidation state oxides are acidic.

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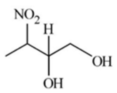
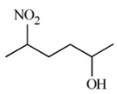
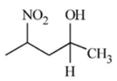
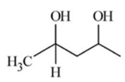
100. Which amongst the following will be most readily dehydrated under acidic conditions ?

(1) 

(2) 

(3) 

(4) 

Correct Answer: (4) Option 4

Solution:

Step 1: Understanding the Question:
The question asks to identify which of the given alcohols will undergo acid-catalyzed dehydration most readily.

Step 2: Key Formula or Approach:
The acid-catalyzed dehydration of alcohols typically proceeds via an E1 mechanism for secondary and tertiary alcohols. The rate-determining step of this mechanism is the formation

of a carbocation intermediate after the protonated hydroxyl group leaves as water. Therefore, the rate of dehydration is directly related to the stability of the carbocation formed. The more stable the carbocation, the faster the reaction.
The stability of carbocations is influenced by:
- **Inductive effect:** Electron-donating groups (EDG) like alkyl groups (+I effect) stabilize carbocations. Electron-withdrawing groups (EWG) like -NO_2 (-I effect) destabilize them.
- **Resonance/Mesomeric effect:** Resonance can delocalize the positive charge, greatly increasing stability. EWGs like -NO_2 also have a strong -R effect, which destabilizes adjacent carbocations.
- **Hyperconjugation:** C-H bonds adjacent to the positively charged carbon can also help delocalize the charge.

Step 3: Detailed Explanation:
Let's analyze the most stable carbocation that can be formed from each alcohol:
- **(1):** This is a diol. Loss of the OH at C-1 would form a secondary carbocation that is severely destabilized by the strongly electron-withdrawing -NO_2 group on the same carbon. Loss of the OH at C-2 would form a primary carbocation. Both are very unstable.
- **(2):** Loss of the OH group forms a secondary carbocation. This carbocation is destabilized by the inductive effect (-I) of the nitrophenyl group, even though it's not directly adjacent.
- **(3):** Loss of the OH group forms a secondary carbocation. It is stabilized by the methyl group (+I effect, hyperconjugation) but strongly destabilized by the adjacent -NO_2 group (-I, -R effects). The destabilizing effect of the nitro group is dominant.
- **(4):** This is a diol (propane-1,2-diol). Loss of the OH group from C-2 would form a secondary carbocation. Loss of the OH group from C-1 would also form a secondary carbocation. Let's consider the loss of OH from C-2. The resulting carbocation at C-2 is stabilized by the adjacent methyl group (+I effect and hyperconjugation). There are no strong electron-withdrawing groups present. This carbocation is the most stable among all the possibilities.

Step 4: Final Answer:
Comparing the stability of the carbocation intermediates, the one formed from alcohol (4) is the most stable due to the stabilizing effect of the methyl group and the absence of any destabilizing electron-withdrawing groups. Therefore, compound (4) will be dehydrated most readily.

Quick Tip

When assessing the rate of reactions that proceed through carbocation intermediates (like S_N1 and E1), always focus on the stability of the carbocation. Look for stabilizing factors (alkyl groups, resonance) and destabilizing factors (electron-withdrawing groups).

Botany

101. In tissue culture experiments, leaf mesophyll cells are put in a culture medium to form callus. This phenomenon may be called as -

(A) Development
(B) Senescence
(C) Differentiation
(D) Dedifferentiation

Correct Answer: (D) Dedifferentiation

Solution:

Step 1: Understanding the Question
The question describes the process in plant tissue culture where specialized cells (leaf mesophyll) are induced to form a mass of unspecialized, dividing cells (callus). We need to identify the correct biological term for this process.

Step 2: Detailed Explanation
Let's define the relevant terms:
Differentiation: The process by which cells become specialized in structure and function (e.g., a meristematic cell becomes a mesophyll cell).
Dedifferentiation: The process by which mature, differentiated cells lose their specialization and revert to a meristematic state, regaining the capacity for cell division. This is exactly what happens when a mesophyll cell (a differentiated cell) is placed on a suitable nutrient medium and forms a callus (an undifferentiated mass of cells).
Redifferentiation: The process where dedifferentiated cells (like those in a callus) divide and differentiate again to form new specialized cells, tissues, and organs, eventually forming a new plantlet.
Senescence: The process of aging in cells or organisms.

The phenomenon described in the question is the reversal of differentiation.

Step 3: Final Answer
The conversion of differentiated mesophyll cells into an undifferentiated callus is known as dedifferentiation. Therefore, option (D) is the correct answer.

> Quick Tip
>
> Memorize the sequence of events in micropropagation via callus formation: 1. **Explant** (differentiated tissue) 2. **Dedifferentiation** $\rightarrow$ **Callus** (undifferentiated) 3. **Redifferentiation** $\rightarrow$ **Plantlet** (differentiated organs)

---

102. In the equation GPP - R = NPP
GPP is Gross Primary Productivity
NPP is Net Primary Productivity

R here is

(A) Respiratory loss
(B) Reproductive allocation
(C) Photosynthetically active radiation
(D) Respiratory quotient

Correct Answer: (A) Respiratory loss

Solution:

Step 1: Understanding the Question
The question asks to identify the term 'R' in the ecological equation relating Gross Primary Productivity (GPP) and Net Primary Productivity (NPP).

Step 2: Key Formula or Approach
The relationship between GPP, NPP, and R is given by the equation:

$$\text{NPP} = \text{GPP} - \text{R}$$

Step 3: Detailed Explanation
Gross Primary Productivity (GPP): This is the rate at which solar energy is captured in sugar molecules during photosynthesis by producers (like plants). It represents the total amount of energy produced.
Net Primary Productivity (NPP): This is the energy that remains as biomass after the producers have met their own energetic needs. It is the energy available to the consumers in the ecosystem.
Producers use a significant portion of the energy they produce (GPP) for their own metabolic activities, primarily cellular respiration. The energy consumed during respiration is lost as heat.
This loss of energy through respiration is represented by 'R'.
So, the equation states that Net Primary Productivity is what's left of the Gross Primary Productivity after the Respiratory losses are accounted for.
Therefore, R stands for Respiratory loss.

Step 4: Final Answer
In the given equation, R represents the energy lost by the producers through respiration. Thus, option (A) is correct.

Quick Tip

Think of GPP as the 'gross income' of an ecosystem and NPP as the 'net income' or 'take-home pay'. The difference, 'R', is the 'expenditure' needed to run the system (i.e., respiration).

103. Given below are two statements :
Statement I: The forces generated by transpiration can lift a xylem-sized column of water over 130 meters height.
Statement II: Transpiration cools leaf surfaces sometimes 10 to 15 degrees, by evaporative cooling.
In the light of the above statements, choose the most appropriate answer from the options given below :

(A) Statement I is correct but Statement II is incorrect.
(B) Statement I is incorrect but Statement II is correct.
(C) Both Statement I and Statement II are correct.
(D) Both Statement I and Statement II are incorrect.

Correct Answer: (C) Both Statement I and Statement II are correct.

Solution:

Step 1: Analyzing the Statements
We need to evaluate the scientific accuracy of both Statement I and Statement II regarding the effects of transpiration.

Step 2: Evaluation of Statement I
This statement refers to the transpiration pull model, also known as the cohesion-tension theory. This theory explains the ascent of sap in tall trees. The key forces involved are:
- **Cohesion:** Strong mutual attraction between water molecules.
- **Adhesion:** Attraction of water molecules to the polar surfaces of xylem vessels.
- **Tension (Transpiration Pull):** A negative pressure potential created in the xylem as water evaporates from the leaves.
These properties give water high tensile strength, allowing it to form an unbroken column that can be pulled up from the roots to the top of the tallest trees. The tallest trees, like the Coast Redwood (Sequoia sempervirens), can exceed 115 meters. The transpiration pull is strong enough to lift water to these heights and even higher (theoretically up to several hundred meters), so lifting it over 130 meters is physically possible. **Therefore, Statement I is correct.**

Step 3: Evaluation of Statement II
This statement describes the cooling effect of transpiration. When water changes from a liquid to a gas (evaporation) on the leaf surface, it absorbs a significant amount of energy from the leaf. This is called the latent heat of vaporization. This loss of heat energy cools the leaf surface, preventing it from getting damaged by high temperatures, especially under intense sunlight. A cooling effect of 10 to 15 degrees Celsius is a well-documented and accepted phenomenon. **Therefore, Statement II is also correct.**

Step 4: Final Answer
Since both Statement I and Statement II are factually correct, the correct option is (C).

Quick Tip

Remember the dual importance of transpiration: it's the 'engine' that pulls water and minerals up the plant (ascent of sap), and it's the plant's 'air conditioner' that prevents leaves from overheating.

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104. In angiosperm, the haploid, diploid and triploid structures of a fertilized embryo sac sequentially are :

(A) Synergids, Zygote and Primary endosperm nucleus
(B) Synergids, antipodals and Polar nuclei
(C) Synergids, Primary endosperm nucleus and zygote
(D) Antipodals, synergids, and primary endosperm nucleus

Correct Answer: (A) Synergids, Zygote and Primary endosperm nucleus

Solution:

Step 1: Understanding the Question
The question asks to identify a sequence of structures from a fertilized embryo sac that are haploid (n), diploid (2n), and triploid (3n), respectively.

Step 2: Detailed Explanation
Let's determine the ploidy level of the key structures in an angiosperm embryo sac after fertilization:
Haploid (n) structures: Before fertilization, the egg cell, synergids, and antipodal cells are all haploid. After fertilization, the synergids and antipodals typically degenerate, but they are still considered haploid structures of the embryo sac.
Diploid (2n) structures: The **zygote** is the primary diploid structure, formed by the fusion of one male gamete (n) with the egg cell (n). It develops into the embryo.
Triploid (3n) structures: The **Primary Endosperm Nucleus (PEN)** is the primary triploid structure. It is formed by the fusion of the second male gamete (n) with the diploid central cell (which contains two polar nuclei, n + n). The PEN develops into the endosperm, which provides nourishment to the developing embryo.

Now let's evaluate the options based on the required sequence (haploid, diploid, triploid):
(A) **Synergids (n), Zygote (2n), Primary endosperm nucleus (3n).** This sequence matches the n, 2n, 3n pattern correctly.
(B) Synergids (n), antipodals (n), Polar nuclei (n+n, so diploid before fusion). This is not the correct sequence.
(C) Synergids (n), Primary endosperm nucleus (3n), zygote (2n). The order is incorrect.
(D) Antipodals (n), synergids (n), primary endosperm nucleus (3n). This does not follow the n, 2n, 3n pattern.

Step 3: Final Answer
The correct sequence of haploid, diploid, and triploid structures is Synergids, Zygote, and Primary endosperm nucleus. Therefore, option (A) is correct.

> Quick Tip
>
> Remember the "double fertilization" process: - 1st fertilization: Male gamete (n) + Egg (n) $\rightarrow$ Zygote (2n) - 2nd fertilization: Male gamete (n) + Central Cell (2n) $\rightarrow$ PEN (3n) This is the source of the diploid and triploid structures. Any remaining cells of the original embryo sac (synergids, antipodals) are haploid.

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105. Given below are two statements :
Statement I: Endarch and exarch are the terms often used for describing the position of secondary xylem in the plant body.
Statement II: Exarch condition is the most common feature of the root system.
In the light of the above statements, choose the correct answer from the options given below :

(A) Statement I is correct but Statement II is false.
(B) Statement I is incorrect but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (B) Statement I is incorrect but Statement II is true.

Solution:

Step 1: Analyzing the Statements
We need to evaluate the correctness of two statements regarding xylem development patterns.

Step 2: Evaluation of Statement I
The terms 'endarch' and 'exarch' describe the pattern of development of **primary xylem**, not secondary xylem.
- **Endarch:** The protoxylem (first-formed primary xylem) is located towards the center (pith), and the metaxylem (later-formed primary xylem) is towards the periphery. This is characteristic of stems.
- **Exarch:** The protoxylem is located towards the periphery, and the metaxylem is towards the center. This is characteristic of roots.
Secondary xylem is formed by the vascular cambium and does not follow these developmental patterns. Therefore, **Statement I is incorrect**.

Step 3: Evaluation of Statement II
This statement claims that the exarch condition is the most common feature of the root system. As explained above, the arrangement of primary xylem in roots where the protoxylem is on the

outer side and metaxylem is on the inner side is called the exarch condition. This is a defining characteristic of roots in dicots and monocots. Therefore, **Statement II is true**.

Step 4: Final Answer
Statement I is incorrect, and Statement II is true. This corresponds to option (B).

> Quick Tip
>
> Remember the association: **Ex**arch is in roots (think **ex**it, outside), and **En**darch is in stems (think **en**ter, inside). These terms apply only to primary xylem.

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106. How many ATP and $NADPH_2$ are required for the synthesis of one molecule of Glucose during Calvin cycle?

(A) 12 ATP and 16 $NADPH_2$
(B) 18 ATP and 16 $NADPH_2$
(C) 12 ATP and 12 $NADPH_2$
(D) 18 ATP and 12 $NADPH_2$

Correct Answer: (D) 18 ATP and 12 $NADPH_2$

Solution:

Step 1: Understanding the Question
The question asks for the total number of ATP and $NADPH_2$ molecules required to produce one molecule of glucose via the Calvin cycle.

Step 2: Key Formula or Approach
The synthesis of one molecule of glucose ($C_6H_{12}O_6$) requires the fixation of 6 molecules of carbon dioxide (CO_2). We need to calculate the energy requirement for 6 turns of the Calvin cycle.

Step 3: Detailed Explanation
The Calvin cycle has three main stages:
1. Carboxylation: 1 CO_2 combines with 1 RuBP. No energy is used here.
2. Reduction: In this stage, the product of carboxylation (2 molecules of 3-PGA) is reduced to 2 molecules of triose phosphate (G3P). This process requires energy for each molecule of CO_2 fixed:
- 2 ATP are used for phosphorylation.
- 2 $NADPH_2$ are used for reduction.
3. Regeneration: To regenerate the initial CO_2 acceptor (RuBP) from the triose phosphates, more energy is required. For each molecule of CO_2 fixed:
- 1 ATP is used.

Calculation for 1 turn (1 CO_2):
- Total ATP used = 2 (from Reduction) + 1 (from Regeneration) = **3 ATP**
- Total $NADPH_2$ used = **2 $NADPH_2$** (from Reduction)

Calculation for 6 turns (to produce 1 Glucose):
To synthesize one molecule of glucose (C_6), the cycle must fix 6 molecules of CO_2. Therefore, we multiply the requirements for one turn by 6:
- Total ATP = 6 turns × 3 ATP/turn = **18 ATP**
- Total $NADPH_2$ = 6 turns × 2 $NADPH_2$/turn = **12 $NADPH_2$**

Step 4: Final Answer
The synthesis of one molecule of glucose requires 18 ATP and 12 $NADPH_2$. Therefore, option (D) is the correct answer.

Quick Tip

For a quick check, remember the ratio for one CO_2 fixed is 3 ATP : 2 $NADPH_2$. To make a 6-carbon glucose, multiply by 6, which gives 18 ATP : 12 $NADPH_2$.

---

107. Spraying of which of the following phytohormone on juvenile conifers helps in hastening the maturity period, that leads to early seed production?

(A) Zeatin
(B) Abscisic Acid
(C) Indole-3-butyric Acid
(D) Gibberellic Acid

Correct Answer: (D) Gibberellic Acid

Solution:

Step 1: Understanding the Question
The question asks which plant hormone is used to speed up the maturation process in juvenile conifers to promote early seed production.

Step 2: Detailed Explanation
Let's examine the functions of the listed phytohormones:
(A) Zeatin: A type of cytokinin, primarily involved in promoting cell division (cytokinesis), chloroplast development, and delaying senescence. It does not primarily hasten maturity.
(B) Abscisic Acid (ABA): Generally considered a plant growth inhibitor. It promotes dormancy, stomatal closure under stress, and senescence. It would be counterproductive for hastening maturity for seed production.
(C) Indole-3-butyric Acid (IBA): A type of auxin, mainly used to promote root formation in plant cuttings. It is involved in cell elongation and apical dominance but not in hastening

the reproductive phase.
(D) Gibberellic Acid (GA): Gibberellins have several roles, including stem elongation (bolting), breaking dormancy, and promoting flowering and fruit development. A key application in horticulture and forestry is spraying juvenile conifers with GAs to overcome the juvenile phase and induce early flowering and seed production. This is commercially important for breeding programs.

Step 3: Final Answer
Gibberellic acid is the phytohormone used to hasten the maturity period in juvenile conifers, leading to early seed production. Hence, option (D) is correct.

Quick Tip

For exams, create a table summarizing the major functions and commercial applications of the five main classes of plant hormones: Auxins, Gibberellins, Cytokinins, Abscisic Acid, and Ethylene. This will help you quickly recall their specific roles.

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108. Among eukaryotes, replication of DNA takes place in -

(A) G_1 phase
(B) G_2 phase
(C) M phase
(D) S phase

Correct Answer: (D) S phase

Solution:

Step 1: Understanding the Question
The question asks to identify the specific phase of the eukaryotic cell cycle during which DNA replication occurs.

Step 2: Detailed Explanation
The eukaryotic cell cycle is divided into two main stages: Interphase and M phase (Mitotic phase). Interphase is further subdivided into three phases:
G_1 phase (Gap 1): This is the phase of cell growth where the cell synthesizes proteins and organelles. The cell is metabolically active, but DNA does not replicate.
S phase (Synthesis phase): This is the specific phase where DNA replication takes place. The amount of DNA in the cell doubles, but the chromosome number remains the same (each chromosome now consists of two sister chromatids).
G_2 phase (Gap 2): The cell continues to grow and synthesizes proteins and organelles in preparation for mitosis.
M phase (Mitotic phase): This is the phase where the cell divides its nucleus (mitosis) and

cytoplasm (cytokinesis) to form two daughter cells.

Step 3: Final Answer
DNA replication is the defining event of the S phase. Therefore, option (D) is the correct answer.

> **Quick Tip**
>
> Remember the cell cycle phases in order: $G_1 \rightarrow S \rightarrow G_2 \rightarrow M$. Associate 'S' with 'Synthesis' of DNA. This is a fundamental concept in cell biology.

---

109. Identify the correct statements :
A. Detrivores perform fragmentation.
B. The humus is further degraded by some microbes during mineralization.
C. Water soluble inorganic nutrients go down into the soil and get precipitated by a process called leaching.
D. The detritus food chain begins with living organisms.
E. Earthworms break down detritus into smaller particles by a process called catabolism.
Choose the correct answer from the options given below :

(A) C, D, E only
(B) D, E, A only
(C) A, B, C only
(D) B, C, D only

Correct Answer: (C) A, B, C only

Solution:

Step 1: Understanding the Question
The question asks to identify the correct statements about the process of decomposition from a given list of five statements.

Step 2: Detailed Explanation
Let's analyze each statement:
A. Detrivores perform fragmentation. This is **correct**. Fragmentation is the physical breakdown of large pieces of dead organic matter (detritus) into smaller particles. This is done by detritivores like earthworms and termites, which increases the surface area for microbial action.
B. The humus is further degraded by some microbes during mineralization. This is **correct**. Humus is a dark, amorphous, colloid-like substance that is highly resistant to microbial action and decomposes very slowly. The slow process by which microbes eventually degrade humus to release inorganic nutrients is called mineralization.
C. Water soluble inorganic nutrients go down into the soil and get precipitated by

a process called leaching. This statement is largely **correct**. Leaching is the process where water-soluble nutrients percolate down through the soil profile and can become unavailable to plants. The term "precipitated" might be slightly imprecise in some contexts, but the core idea of nutrients moving down into the soil is correct.
D. The detritus food chain begins with living organisms. This is **incorrect**. The detritus food chain (DFC) begins with dead organic matter (detritus). The grazing food chain (GFC) begins with living organisms (producers).
E. Earthworms break down detritus into smaller particles by a process called catabolism. This is **incorrect**. Earthworms break down detritus by **fragmentation**. Catabolism is the enzymatic breakdown of complex organic matter into simpler inorganic substances by decomposer organisms like bacteria and fungi.

Step 3: Final Answer
Statements A, B, and C are correct, while D and E are incorrect. Therefore, the correct option is (C) A, B, C only.

Quick Tip

Remember the key steps in decomposition: 1. **Fragmentation** (physical breakdown by detritivores). 2. **Leaching** (water-soluble nutrients move down). 3. **Catabolism** (enzymatic breakdown by microbes). 4. **Humification** (formation of humus). 5. **Mineralization** (release of inorganic nutrients from humus).

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110. Identify the pair of heterosporous pteridophytes among the following :

(A) Psilotum and Salvinia
(B) Equisetum and Salvinia
(C) Lycopodium and Selaginella
(D) Selaginella and Salvinia

Correct Answer: (D) Selaginella and Salvinia

Solution:

Step 1: Understanding the Question
The question requires identifying a pair of pteridophytes that are both heterosporous.

Step 2: Detailed Explanation
Pteridophytes are classified based on the types of spores they produce:
Homosporous Pteridophytes: These plants produce only one type of spore, which grows into a bisexual (monoecious) gametophyte that bears both antheridia and archegonia. The majority of pteridophytes are homosporous. Examples include Psilotum, Lycopodium, and Equisetum.
Heterosporous Pteridophytes: These plants produce two distinct types of spores: smaller

microspores and larger megaspores. Microspores germinate to form male gametophytes, and megaspores germinate to form female gametophytes. This condition is a precursor to the seed habit seen in gymnosperms and angiosperms. Key examples of heterosporous pteridophytes are Selaginella, Salvinia, Azolla, and Marsilea.

Analyzing the options:
(A) Psilotum is homosporous, Salvinia is heterosporous.
(B) Equisetum is homosporous, Salvinia is heterosporous.
(C) Lycopodium is homosporous, Selaginella is heterosporous.
(D) Selaginella is heterosporous, and Salvinia is also heterosporous.

Step 3: Final Answer
Both Selaginella and Salvinia are examples of heterosporous pteridophytes. Therefore, option (D) is the correct pair.

> **Quick Tip**
>
> For exams, remember the main examples for each category. A good mnemonic for heterosporous pteridophytes is "Selaginella, Salvinia, Azolla, Marsilea". Most other common pteridophytes like Pteris, Dryopteris, Equisetum, and Lycopodium are homosporous.

---

111. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R :
Assertion A: ATP is used at two steps in glycolysis.
Reason R: First ATP is used in converting glucose into glucose-6-phosphate and second ATP is used in conversion of fructose-6-phosphate into fructose-1-6-diphosphate.
In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (C) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Understanding the Statements
The question presents an Assertion (A) stating that ATP is consumed at two points in glycolysis, and a Reason (R) that specifies these two points. We must evaluate their accuracy and relationship.

Step 2: Detailed Explanation
Assertion A: ATP is used at two steps in glycolysis.
This statement is **true**. Glycolysis is divided into a preparatory (investment) phase and a payoff phase. During the preparatory phase, the cell invests two molecules of ATP to activate the glucose molecule.

Reason R: First ATP is used in converting glucose into glucose-6-phosphate and second ATP is used in conversion of fructose-6-phosphate into fructose-1-6-diphosphate.
Let's analyze the two parts of this reason: 1. The first ATP is indeed used to phosphorylate glucose to glucose-6-phosphate, a reaction catalyzed by hexokinase. This part is correct. 2. The second ATP is used to phosphorylate fructose-6-phosphate to fructose-1,6-bisphosphate, catalyzed by phosphofructokinase.
Connecting A and R:
Since Assertion A is true but Reason R is also true due the correct choice is that A is true but R is the correct explanation of A.

> **Quick Tip**
>
> Pay close attention to biochemical terminology. The difference between "bisphosphate" (two separate phosphate groups) and "diphosphate" (two linked phosphate groups, like in ADP) can be the key to a correct answer in questions on metabolic pathways.

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112. Among 'The Evil Quartet', which one is considered the most important cause driving extinction of species?

(A) Alien species invasions
(B) Co-extinctions
(C) Habitat loss and fragmentation
(D) Over exploitation for economic gain

Correct Answer: (C) Habitat loss and fragmentation

Solution:

Step 1: Understanding the Question
The question asks to identify the most significant driver of species extinction among the four major causes collectively known as 'The Evil Quartet'.

Step 2: Detailed Explanation
'The Evil Quartet' is a term used to describe the four main human-caused drivers of biodiversity loss:
1. **Habitat loss and fragmentation:** This involves the destruction or division of natural habitats (e.g., deforestation, urbanization, conversion to agriculture). It is widely regarded by ecologists as the single greatest threat to biodiversity globally. When an organism's home is

destroyed or broken into small, isolated patches, its populations decline, and it becomes vulnerable to extinction.
2. **Over-exploitation:** This refers to harvesting species from the wild at rates faster than natural populations can recover (e.g., overfishing, overhunting). It is a major threat to many large animals and commercially valuable species.
3. **Alien species invasions:** The introduction of non-native species into an ecosystem can disrupt food webs, outcompete native species for resources, and introduce diseases, leading to the decline and extinction of native species.
4. **Co-extinctions:** This occurs when the extinction of one species causes the extinction of another species that depended on it (e.g., a specialist parasite losing its only host).

While all four are serious threats, habitat loss and fragmentation affects the largest number of species across all taxa and is the primary cause of the current extinction crisis.

Step 3: Final Answer
Among the four major causes of biodiversity loss, habitat loss and fragmentation is considered the most important driver of species extinction. Thus, option (C) is correct.

> **Quick Tip**
> When thinking about threats to biodiversity, always consider habitat loss as the number one factor. The other threats often exacerbate the problem created by a shrinking and fragmented habitat.

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113. Unequivocal proof that DNA is the genetic material was first proposed by

(A) Avery, Macleoid and McCarthy
(B) Wilkins and Franklin
(C) Frederick Griffith
(D) Alfred Hershey and Martha Chase

Correct Answer: (D) Alfred Hershey and Martha Chase

Solution:

Step 1: Understanding the Question
The question asks to identify the scientists who provided the definitive or "unequivocal" proof that DNA is the substance of heredity.

Step 2: Detailed Explanation
The discovery of DNA as the genetic material was a gradual process with contributions from several scientists:
(C) Frederick Griffith (1928): His "transforming principle" experiment with *Streptococcus pneumoniae* showed that a substance from dead pathogenic bacteria could transform living

non-pathogenic bacteria, but he did not identify the substance.
(A) Avery, Macleod, and McCarthy (1944): They expanded on Griffith's work and showed through systematic experiments using enzymes (proteases, RNases, DNases) that the transforming principle was DNA. This was strong evidence, but not universally accepted, as some scientists still believed that protein contaminants might be responsible.
(D) Alfred Hershey and Martha Chase (1952): They conducted the famous "blender experiment" using bacteriophages (viruses that infect bacteria). They used radioactive isotopes to label the two main components of the virus:
- **Protein coat** was labeled with radioactive sulfur (^{35}S), as sulfur is present in proteins but not DNA.
- **DNA core** was labeled with radioactive phosphorus (^{32}P), as phosphorus is present in DNA but not proteins.

They found that only the radioactive phosphorus (^{32}P) entered the bacterial cells, while the radioactive sulfur (^{35}S) remained outside. Since the bacteria produced new viruses, it was the DNA that must have carried the genetic information. This experiment provided the clear, unambiguous, and unequivocal proof that DNA is the genetic material.
(B) Wilkins and Franklin: They were instrumental in determining the structure of DNA using X-ray diffraction, which was crucial for Watson and Crick's model, but they did not conduct experiments to prove it was the genetic material.

Step 3: Final Answer
The Hershey-Chase experiment is considered the conclusive evidence that established DNA as the genetic material. Therefore, option (D) is correct.

> Quick Tip
>
> Remember the progression of discovery: Griffith showed *something* transformed bacteria. Avery et al. showed it was *probably* DNA. Hershey and Chase showed it was *definitively* DNA by tracking radio-labeled components.

---

114. The thickness of ozone in a column of air in the atmosphere is measured in terms of :

(A) Decameter
(B) Kilobase
(C) Dobson units
(D) Decibels

Correct Answer: (C) Dobson units

Solution:

Step 1: Understanding the Question
The question asks for the standard unit of measurement for the thickness of the atmospheric

ozone layer.

Step 2: Detailed Explanation
Let's analyze the given units:
(A) Decameter: A unit of length, equal to 10 meters. It is not used for measuring atmospheric gases.
(B) Kilobase (kb): A unit used in molecular biology to measure the length of DNA or RNA molecules, equal to 1000 base pairs. It is irrelevant to atmospheric science.
(C) Dobson Units (DU): This is the standard unit used to measure the total amount of ozone in a vertical column of the atmosphere (the ozone column). One Dobson Unit is defined as the thickness (in units of 10 µm) of the layer of pure ozone that would be formed if all the ozone in the column were compressed to standard temperature (0°C) and pressure (1 atm). For example, 300 DU corresponds to a 3 mm thick layer of pure ozone.
(D) Decibels (dB): A logarithmic unit used to measure the intensity of sound or the power level of an electrical signal. It is unrelated to ozone measurement.

Step 3: Final Answer
The correct unit for measuring the thickness of the ozone layer is the Dobson unit. Therefore, option (C) is the correct answer.

> **Quick Tip**
>
> For environmental science topics, create a list of important quantities and their specific units of measurement. For instance: Ozone layer - Dobson Units; Sound intensity - Decibels; Radioactivity - Becquerel/Curie.

---

115. Which micronutrient is required for splitting of water molecule during photosynthesis?

(A) magnesium
(B) copper
(C) manganese
(D) molybdenum

Correct Answer: (C) manganese

Solution:

Step 1: Understanding the Question
The question asks to identify the micronutrient that plays a crucial role in the photolysis (splitting) of water during the light-dependent reactions of photosynthesis.

Step 2: Detailed Explanation
The splitting of water molecules occurs in Photosystem II (PS II) and is catalyzed by a protein

complex called the Oxygen-Evolving Complex (OEC). The reaction is:

$$2H_2O \rightarrow 4H^+ + O_2 + 4e^-$$

This process requires the presence of specific inorganic ions as cofactors. The core of the OEC contains a cluster of four **manganese (Mn)** ions and one calcium (Ca^{2+}) ion. Chloride ions (Cl^-) are also essential for this process.
Let's look at the roles of the other options:
(A) Magnesium (Mg): It is the central atom in the chlorophyll molecule, essential for capturing light energy, but not for splitting water.
(B) Copper (Cu): It is a component of plastocyanin, an electron carrier protein that transfers electrons from cytochrome b6f complex to PS I.
(D) Molybdenum (Mo): It is a component of enzymes like nitrate reductase and nitrogenase, involved in nitrogen metabolism.

Step 3: Final Answer
Manganese (Mn) is the essential micronutrient required for the enzymatic splitting of water in photosynthesis. Therefore, option (C) is correct.

Quick Tip

Create a list matching essential micronutrients to their key functions in plants. For example: Mn $\rightarrow$ Photolysis of water; Zn $\rightarrow$ Auxin synthesis; Mo $\rightarrow$ Nitrogen fixation; B $\rightarrow$ Pollen germination. This is a high-yield topic for exams.

---

116. Family Fabaceae differs from Solanaceae and Liliaceae. With respect to the stamens, pick out the characteristics specific to family Fabaceae but not found in Solanaceae or Liliaceae.

(A) Monoadelphous and Monothecous anthers
(B) Epiphyllous and Dithecous anthers
(C) Diadelphous and Dithecous anthers
(D) Polyadelphous and epipetalous stamens

Correct Answer: (C) Diadelphous and Dithecous anthers

Solution:

Step 1: Understanding the Question
The question asks to identify the specific characteristics of the androecium (stamens) that are unique to the family Fabaceae when compared to Solanaceae and Liliaceae.

Step 2: Detailed Explanation
Let's analyze the stamen characteristics of each family:

Family Fabaceae (Subfamily Papilionoideae): The androecium typically consists of ten stamens. A key feature is that they are **diadelphous**, meaning they are united into two bundles. The common arrangement is (9)+1, where nine stamens are fused to form a tube and one is free. The anthers are **dithecous** (having two lobes).
Family Solanaceae: The androecium has five stamens which are **epipetalous** (attached to the petals). They are not fused into bundles.
Family Liliaceae: The androecium has six stamens, arranged in two whorls of three (3+3). They are often **epiphyllous** or **epitepalous** (attached to the tepals).

Now let's evaluate the options:
(A) Monoadelphous (stamens in one bundle) is characteristic of Malvaceae.
(B) Epiphyllous condition is characteristic of Liliaceae.
(C) Diadelphous condition is a hallmark of Fabaceae and is not found in Solanaceae or Liliaceae. Dithecous anthers are common, but the diadelphous arrangement is the distinguishing feature.
(D) Polyadelphous (stamens in many bundles) is found in families like Rutaceae (e.g., Citrus). Epipetalous condition is found in Solanaceae.

Step 3: Final Answer
The diadelphous condition of stamens is a specific characteristic of family Fabaceae that distinguishes it from Solanaceae and Liliaceae. Thus, option (C) is the correct answer.

Quick Tip

When studying plant families, focus on unique floral characteristics. For Fabaceae, remember the vexillary aestivation of the corolla and the diadelphous (9)+1 arrangement of stamens. These are frequently tested features.

---

117. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R :
Assertion A: The first stage of gametophyte in the life cycle of moss is protonema stage.
Reason R: Protonema develops directly from spores produced in capsule.
In the light of the above statements, choose the most appropriate answer from the options given below :

(A) A is correct but R is not correct.
(B) A is not correct but R is correct.
(C) Both A and R are correct and R is the correct explanation of A.
(D) Both A and R are correct but R is NOT the correct explanation of A.

Correct Answer: (C) Both A and R are correct and R is the correct explanation of A.

Solution:

Step 1: Understanding the Statements

The question presents an Assertion (A) about the protonema being the first stage of the moss gametophyte and a Reason (R) explaining its origin from a spore. We need to assess the truthfulness of both and their relationship.

Step 2: Detailed Explanation

Moss Life Cycle: The life cycle of a moss (a bryophyte) involves an alternation of generations between a haploid gametophyte and a diploid sporophyte.

1. The diploid sporophyte, which remains attached to the gametophyte, has a structure called a capsule. Meiosis occurs within the capsule to produce haploid spores.
2. When a haploid spore lands on a suitable substrate, it germinates.
3. The spore develops into a filamentous, green, creeping structure called the **protonema**. This is the juvenile gametophyte stage. So, Reason R, which states that the protonema develops directly from spores produced in the capsule, is correct.
4. The protonema is indeed the very first stage of the gametophyte generation. Later, leafy buds arise from this protonema, which develop into the mature, upright, leafy gametophyte (the main moss plant). Therefore, Assertion A is also correct.

Connecting A and R: The reason the protonema is the first stage of the gametophyte (Assertion A) is precisely because it is the structure that directly emerges from the germinating spore (Reason R). The spore is the beginning of the gametophytic generation. Thus, Reason R provides the correct explanation for Assertion A.

Step 3: Final Answer

Both Assertion A and Reason R are correct statements, and Reason R is the correct explanation for Assertion A. Therefore, option (C) is the correct answer.

> Quick Tip
>
> To master plant life cycles, draw them out. For mosses, remember the sequence: Spore (n) $\rightarrow$ Protonema (n) $\rightarrow$ Leafy Gametophyte (n) $\rightarrow$ Gametes (n) $\rightarrow$ Zygote (2n) $\rightarrow$ Sporophyte (2n) $\rightarrow$ Spore (n). This visual map helps in answering any related question.

---

118. Expressed Sequence Tags (ESTs) refers to

(A) All genes whether expressed or unexpressed.
(B) Certain important expressed genes.
(C) All genes that are expressed as RNA.
(D) All genes that are expressed as proteins.

Correct Answer: (C) All genes that are expressed as RNA.

Solution:

Step 1: Understanding the Question
The question asks for the definition of Expressed Sequence Tags (ESTs).

Step 2: Detailed Explanation
Expressed Sequence Tags (ESTs) are short, single-pass sequence reads derived from Complementary DNA (cDNA) libraries.
1. To generate a cDNA library, messenger RNA (mRNA) is first isolated from a cell. mRNA represents the genes that are actively being transcribed, or "expressed."
2. The enzyme reverse transcriptase is used to synthesize a single-stranded DNA molecule complementary to the mRNA template. This is cDNA.
3. Short fragments of these cDNAs are sequenced to generate ESTs.
Therefore, ESTs are tags or identifiers for genes that are expressed as RNA (specifically, mRNA) in a given tissue at a given time.

Analyzing the options:
(A) This is incorrect. ESTs only represent expressed genes, not unexpressed ones.
(B) This is incorrect. ESTs are generated from the entire mRNA population, not just "certain important" genes.
(C) This is the most accurate description. ESTs represent all genes that are transcribed into RNA.
(D) This is incorrect. While many RNAs are translated into proteins, not all are (e.g., non-coding RNAs). ESTs represent the RNA transcript itself, not the final protein product.

Step 3: Final Answer
ESTs are sequences derived from mRNA, so they represent all genes that are expressed as RNA. Thus, option (C) is correct.

Quick Tip

Remember that ESTs come from cDNA, and cDNA is made from mRNA. The 'E' in EST stands for "Expressed," which means the gene is transcribed into RNA. This directly links ESTs to genes expressed as RNA.

---

119. Frequency of recombination between gene pairs on same chromosome as a measure of the distance between genes to map their position on chromosome, was used for the first time by

(A) Alfred Sturtevant
(B) Henking
(C) Thomas Hunt Morgan
(D) Sutton and Boveri

Correct Answer: (A) Alfred Sturtevant

Solution:

Step 1: Understanding the Question
The question asks to identify the scientist who first proposed and used the frequency of genetic recombination to determine the relative distances between genes on a chromosome, a technique known as gene mapping.

Step 2: Reviewing the Contributions of the Scientists
(D) Sutton and Boveri: They independently proposed the **Chromosomal Theory of Inheritance** around 1902, stating that genes are located on chromosomes.
(B) Henking: In 1891, while studying spermatogenesis, he observed a specific nuclear structure which he named the X body. This was later identified as the X chromosome.
(C) Thomas Hunt Morgan: Working with fruit flies (Drosophila melanogaster), he provided experimental proof for the Chromosomal Theory of Inheritance. He discovered concepts like linkage (genes on the same chromosome tend to be inherited together) and recombination (crossing over can break linkages). His work laid the foundation for gene mapping.
(A) Alfred Sturtevant: He was a student in T.H. Morgan's lab. In 1913, he had the crucial insight that the **frequency** of recombination between linked genes was a function of their physical distance from each other on the chromosome. He used recombination data to construct the very first **genetic linkage map**.

Step 3: Final Answer
While Morgan's work was foundational, it was his student, Alfred Sturtevant, who first used recombination frequencies as a measure of distance to create a gene map. Therefore, option (A) is the correct answer.

Quick Tip

Associate the scientists with their key "firsts": - **Sutton & Boveri:** First to propose the Chromosomal Theory. - **Morgan:** First to experimentally prove linkage and recombination in Drosophila. - **Sturtevant:** First to create a genetic map using recombination frequency.

---

120. Large, colourful, fragrant flowers with nectar are seen in :

(A) bat pollinated plants
(B) wind pollinated plants
(C) insect pollinated plants
(D) bird pollinated plants

Correct Answer: (C) insect pollinated plants

Solution:

Step 1: Understanding the Question
The question describes a set of floral characteristics (large size, colourful, fragrant, nectar-producing) and asks to identify the corresponding pollination agent. This relates to the concept of pollination syndromes.

Step 2: Detailed Explanation
Different pollination agents are attracted by different floral features:
(A) Bat pollinated plants (Chiropterophily): Flowers are typically large but pale or dull-coloured, open at night, and emit a strong, musty or fermented fruit-like odour. They produce copious nectar.
(B) Wind pollinated plants (Anemophily): Flowers are small, inconspicuous, lack colour, fragrance, and nectar. They are adapted to produce and release large quantities of lightweight pollen.
(C) Insect pollinated plants (Entomophily): Flowers have evolved to attract insects. The combination of being **large** (to be visible), **colourful** (to attract visually), **fragrant** (to attract by smell), and producing **nectar** (as a food reward) is the classic suite of adaptations for insect pollination.
(D) Bird pollinated plants (Ornithophily): Flowers are often large and brightly coloured (especially red and orange), but they typically lack a strong scent as birds have a poor sense of smell. They produce abundant, dilute nectar.

The combination of all four traits listed in the question is most characteristic of insect-pollinated plants.

Step 3: Final Answer
The set of features described perfectly matches the adaptations for attracting insects. Therefore, option (C) is the correct answer.

> Quick Tip
>
> Create a table to remember pollination syndromes. Key columns should be: Pollinator (Wind, Insect, Bird, Bat), Flower Size, Colour, Scent, and Nectar. This helps in quickly answering such questions.

---

121. Movement and accumulation of ions across a membrane against their concentration gradient can be explained by

(A) Passive Transport
(B) Active Transport
(C) Osmosis
(D) Facilitated Diffusion

Correct Answer: (B) Active Transport

Solution:

Step 1: Understanding the Question
The key phrase in the question is "against their concentration gradient". This means moving substances from an area of lower concentration to an area of higher concentration. We need to identify the transport mechanism that allows this.

Step 2: Detailed Explanation
Let's define the different types of membrane transport:
(A) Passive Transport: The movement of substances across a membrane **down** the concentration gradient (from high to low concentration). It does not require metabolic energy.
(C) Osmosis: A specific type of passive transport involving the movement of water across a semipermeable membrane down its water potential gradient.
(D) Facilitated Diffusion: A type of passive transport where substances move down the concentration gradient with the help of membrane proteins (channels or carriers). It does not require energy.
(B) Active Transport: The movement of substances across a membrane **against** their concentration gradient (from low to high concentration). This process is like moving something "uphill" and requires carrier proteins and the expenditure of metabolic energy, typically in the form of ATP.

The accumulation of ions against a gradient is a hallmark of active transport.

Step 3: Final Answer
The movement of ions against a concentration gradient requires energy and is defined as Active Transport. Therefore, option (B) is the correct answer.

> **Quick Tip**
>
> Remember the key distinction: **Passive** = Downhill, no energy. **Active** = Uphill, requires energy (ATP). The phrase "against the concentration gradient" is a direct indicator of active transport.

---

122. Upon exposure to UV radiation, DNA stained with ethidium bromide will show

(A) Bright yellow colour
(B) Bright orange colour
(C) Bright red colour
(D) Bright blue colour

Correct Answer: (B) Bright orange colour

Solution:

Step 1: Understanding the Question
The question asks for the color of fluorescence observed when DNA stained with ethidium bromide (EtBr) is exposed to ultraviolet (UV) light. This is a standard technique in molecular biology.

Step 2: Detailed Explanation
Agarose Gel Electrophoresis: This technique is used to separate DNA fragments based on their size.
Staining with Ethidium Bromide (EtBr): After electrophoresis, the DNA in the gel is invisible. To visualize it, the gel is soaked in a solution containing ethidium bromide. EtBr is a fluorescent dye that intercalates, or inserts itself, between the base pairs of the DNA double helix.
Visualization under UV Light: The EtBr-DNA complex absorbs UV radiation at a wavelength of around 300-360 nm. Upon absorption of this energy, it fluoresces, emitting light in the visible spectrum. The emitted light has a longer wavelength, which appears as a characteristic bright orange colour. This allows the separated DNA bands to be seen and photographed.

Step 3: Final Answer
DNA stained with ethidium bromide fluoresces bright orange when exposed to UV radiation. Thus, option (B) is the correct answer.

> **Quick Tip**
>
> Associate specific stains with their resulting colors in molecular biology techniques. For example, Ethidium Bromide + DNA + UV light = Bright Orange. This is a very common and frequently asked question from biotechnology.

---

123. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R :
Assertion A: Late wood has fewer xylary elements with narrow vessels.
Reason R: Cambium is less active in winters.
In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (C) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Understanding the Statements
The question presents an Assertion (A) about the characteristics of late wood and a Reason (R) explaining the activity of cambium in winter. We need to evaluate if both statements are true and if the reason correctly explains the assertion.

Step 2: Detailed Explanation
Assertion A: Late wood, also known as autumn wood, is formed during the later part of the growing season (autumn/winter). During this period, the environmental conditions are less favourable for growth. This results in the formation of fewer xylary elements, and the vessels (or tracheids) that are formed are narrower in diameter. So, Assertion A is true.

Reason R: The cambium is a layer of actively dividing cells responsible for secondary growth (increase in girth). Its activity is influenced by physiological and environmental factors, such as temperature and water availability. In winters, the conditions are harsh, and the cambium becomes less active. This reduced activity leads to the formation of late wood. So, Reason R is also true.

Connecting A and R: Because the cambium is less active in winters (Reason R), it produces fewer xylem elements with narrow vessels, which is the characteristic feature of late wood (Assertion A). Therefore, Reason R is the correct explanation for Assertion A.

Step 3: Final Answer
Both Assertion A and Reason R are true, and Reason R correctly explains Assertion A. Thus, option (C) is the correct answer.

> Quick Tip
>
> For Assertion-Reason questions, first check the validity of each statement independently. Then, check if the Reason provides a direct explanation for the Assertion by asking "Why?" about the Assertion. If the Reason answers "Why?", it is the correct explanation.

---

124. Axile placentation is observed in

(A) Tomato, Dianthus and Pea
(B) China rose, Petunia and Lemon
(C) Mustard, Cucumber and Primrose
(D) China rose, Beans and Lupin

Correct Answer: (B) China rose, Petunia and Lemon

Solution:

Step 1: Understanding the Question
The question asks to identify the group of plants from the given options that all exhibit axile placentation. Placentation refers to the arrangement of ovules within the ovary.

Step 2: Defining Axile Placentation
In axile placentation, the ovary is partitioned by septa into two or more chambers (locules). The placenta is located on the central axis where the septa meet, and the ovules are attached to this central axis within each locule.

Step 3: Analyzing the Options
Let's determine the type of placentation for each plant in the options:
(A) Tomato, Dianthus and Pea:
- Tomato: Axile placentation.
- Dianthus: Free-central placentation.
- Pea: Marginal placentation.
This option is incorrect as it contains three different types.

(B) China rose, Petunia and Lemon:
- China rose (Hibiscus): Axile placentation.
- Petunia (from Solanaceae family): Axile placentation.
- Lemon (Citrus): Axile placentation.
This option is correct as all three plants show axile placentation.

(C) Mustard, Cucumber and Primrose:
- Mustard: Parietal placentation.
- Cucumber: Parietal placentation.
- Primrose: Free-central placentation.
This option is incorrect.

(D) China rose, Beans and Lupin:
- China rose: Axile placentation.
- Beans (like Pea): Marginal placentation.
- Lupin: Marginal placentation.
This option is incorrect.

Step 4: Final Answer
The only group where all plants listed exhibit axile placentation is China rose, Petunia, and Lemon. Therefore, option (B) is the correct answer.

Quick Tip

To master placentation, associate each type with a common fruit or vegetable you can visualize: - **Axile:** Sliced tomato or lemon. - **Marginal:** Pea pod. - **Parietal:** Sliced cucumber or papaya. - **Free-central:** Primrose flower. - **Basal:** Sunflower seed in the head.

125. The phenomenon of pleiotropism refers to

(A) a single gene affecting multiple phenotypic expression.
(B) more than two genes affecting a single character.
(C) presence of several alleles of a single gene controlling a single crossover.
(D) presence of two alleles, each of the two genes controlling a single trait.

Correct Answer: (A) a single gene affecting multiple phenotypic expression.

Solution:

Step 1: Understanding the Question
The question asks for the definition of pleiotropism (or pleiotropy) in genetics.

Step 2: Detailed Explanation
Let's define the terms in the options:
(A) a single gene affecting multiple phenotypic expression: This is the definition of **pleiotropy**. A single gene can influence several different, often unrelated, traits. A classic example is the gene for phenylketonuria (PKU), which causes mental retardation, reduced hair pigmentation, and skin pigmentation.
(B) more than two genes affecting a single character: This describes **polygenic inheritance**. Traits like height, skin color, and weight are controlled by the cumulative effect of many genes.
(C) presence of several alleles of a single gene controlling a single crossover: This option is confusing and does not describe a standard genetic phenomenon. The presence of several alleles of a single gene is called **multiple allelism**. Crossover is a separate process.
(D) presence of two alleles, each of the two genes controlling a single trait: This is also an incorrectly phrased option. It seems to mix ideas of alleles and multiple genes.

Based on standard genetic definitions, pleiotropy is when one gene influences multiple traits.

Step 3: Final Answer
The correct definition of pleiotropism is a single gene affecting multiple phenotypic expressions. Therefore, option (A) is the correct answer.

> Quick Tip
>
> To avoid confusion, clearly distinguish between pleiotropy and polygenic inheritance.
> **Pleiotropy:** One gene $\rightarrow$ Many traits.
> **Polygenic Inheritance:** Many genes $\rightarrow$ One trait.

126. During the purification process for recombinant DNA technology, addition of

chilled ethanol precipitates out

(A) Histones
(B) Polysaccharides
(C) RNA
(D) DNA

Correct Answer: (D) DNA

Solution:

Step 1: Understanding the Question
The question asks which biological macromolecule is precipitated out from a solution by adding chilled ethanol during the purification step in recombinant DNA technology. This is a standard step in DNA isolation protocols.

Step 2: Detailed Explanation
The process of isolating DNA from cells involves several steps:
1. **Lysis:** Breaking open the cells to release their contents, including DNA, RNA, proteins, and lipids.
2. **Removal of other macromolecules:** Enzymes like proteases (to digest proteins like histones) and RNase (to digest RNA) are often added.
3. **Precipitation of DNA:** After other major contaminants are removed, the DNA is in an aqueous solution. DNA is insoluble in alcohols like ethanol or isopropanol. When chilled ethanol is added to the aqueous solution, the DNA precipitates out as a fine, white, thread-like mass. The low temperature reduces the solubility further, enhancing precipitation. Other soluble components like salts and sugars remain in the solution.

Therefore, the addition of chilled ethanol is the specific step used to precipitate and collect the purified DNA.

Step 3: Final Answer
Based on the principles of DNA isolation, chilled ethanol is used to precipitate DNA from the solution. Hence, option (D) is the correct answer.

> **Quick Tip**
>
> Remember that DNA is insoluble in alcohol. This is a fundamental principle used in almost all DNA extraction kits and protocols. The precipitated DNA can often be visualized as a cloudy or thread-like substance and can be spooled onto a glass rod.

---

127. Which hormone promotes internode/petiole elongation in deep water rice?

(A) Ethylene
(B) 2, 4-D
(C) GA_3
(D) Kinetin

Correct Answer: (A) Ethylene

Solution:

Step 1: Understanding the Question
The question asks to identify the plant hormone responsible for promoting the rapid elongation of internodes or petioles in deep-water rice varieties when they are submerged.

Step 2: Detailed Explanation
Deep-water rice has a remarkable adaptation to survive flooding. When the plant is submerged, the gaseous hormone **ethylene** accumulates in the submerged plant parts because its diffusion out of the plant is blocked by water.
This increased concentration of ethylene triggers a physiological response. It enhances the sensitivity of the cells to gibberellic acid (GA) or promotes GA synthesis, which in turn causes rapid cell division and elongation in the internodes and petioles. This allows the leaves to quickly reach the water surface to continue photosynthesis and gas exchange.
While GA_3 (a gibberellin) is the hormone that directly causes the elongation, ethylene is the primary signal or promoter of this response specifically in the context of submergence in deep-water rice. Given the options, ethylene is the most accurate answer for the hormone that "promotes" this specific adaptive phenomenon.

Step 3: Final Answer
Ethylene is the key hormone that accumulates during submergence and promotes the elongation response in deep-water rice. Thus, option (A) is correct.

Quick Tip

Remember ethylene as the "stress" hormone in some contexts. In deep-water rice, the stress of submergence leads to ethylene accumulation, which triggers the escape mechanism (rapid elongation).

---

128. The process of appearance of recombination nodules occurs at which sub stage of prophase I in meiosis?

(A) Diplotene
(B) Diakinesis
(C) Zygotene
(D) Pachytene

Correct Answer: (D) Pachytene

Solution:

Step 1: Understanding the Question
The question asks to identify the specific substage of Prophase I of meiosis where recombination nodules are observed.

Step 2: Detailed Explanation
Prophase I is the longest phase of meiosis and is divided into five substages:
1. **Leptotene:** Chromosomes start to condense and become visible.
2. **Zygotene:** Homologous chromosomes pair up in a process called synapsis, forming structures called bivalents.
3. **Pachytene:** This is a relatively long stage where the paired homologous chromosomes (bivalents) are clearly visible as tetrads. During this stage, **crossing over** occurs. This is the exchange of genetic material between non-sister chromatids of homologous chromosomes. The sites where crossing over occurs are marked by the appearance of protein complexes called **recombination nodules**.
4. **Diplotene:** The synaptonemal complex dissolves, and the homologous chromosomes start to separate, except at the sites of crossing over. These X-shaped structures are called chiasmata.
5. **Diakinesis:** Chromosomes become fully condensed, and the chiasmata terminalize. The nuclear envelope breaks down.

The appearance of recombination nodules is directly associated with the process of crossing over, which is the characteristic event of the pachytene stage.

Step 3: Final Answer
Recombination nodules, the sites of crossing over, appear during the Pachytene substage of Prophase I. Therefore, option (D) is correct.

> Quick Tip
>
> Use the mnemonic "Lazy Zebra Pushes Down Donkey" to remember the order of Prophase I substages: Leptotene, Zygotene, Pachytene, Diplotene, Diakinesis. Associate **P**achytene with **P**airing and crossing over.

---

129. What is the role of RNA polymerase III in the process of transcription in Eukaryotes?

(A) Transcription of precursor of mRNA
(B) Transcription of only snRNAs
(C) Transcription of rRNAs (28S, 18S and 5.8S)
(D) Transcription of tRNA, 5S rRNA and snRNA

Correct Answer: (D) Transcription of tRNA, 5S rRNA and snRNA

Solution:

Step 1: Understanding the Question
The question asks to identify the specific types of RNA molecules that are transcribed by the enzyme RNA polymerase III in eukaryotic cells.

Step 2: Detailed Explanation
In eukaryotes, there are three main types of RNA polymerases, each with distinct roles in transcription:
RNA Polymerase I: Located in the nucleolus, it is responsible for transcribing most of the ribosomal RNA (rRNA) genes, specifically the 28S, 18S, and 5.8S rRNA molecules. This eliminates option (C).
RNA Polymerase II: Located in the nucleoplasm, its primary role is to transcribe the genes that code for proteins. This means it synthesizes the precursors of messenger RNA (pre-mRNA or hnRNA) and also most small nuclear RNAs (snRNAs). This eliminates option (A).
RNA Polymerase III: Also located in the nucleoplasm, it transcribes genes for smaller RNA molecules. Its main products are transfer RNA (tRNA), 5S ribosomal RNA (5S rRNA), and some small nuclear RNAs (snRNAs, such as U6 snRNA).

Comparing these roles with the given options:
(A) Precursor of mRNA is transcribed by RNA Pol II.
(B) "Only snRNAs" is incorrect; RNA Pol II also transcribes snRNAs, and RNA Pol III transcribes more than just snRNAs.
(C) 28S, 18S, and 5.8S rRNAs are transcribed by RNA Pol I.
(D) tRNA, 5S rRNA, and snRNA are the correct products of RNA Polymerase III.

Step 3: Final Answer
The role of RNA polymerase III is the transcription of tRNA, 5S rRNA, and some snRNAs. Therefore, option (D) is the correct answer.

Quick Tip

Use the mnemonic "1-2-3, R-M-T" to remember the main products of the eukaryotic RNA polymerases: Pol **I** -¿ **r**RNA, Pol **II** -¿ **m**RNA, Pol **III** -¿ **t**RNA. This helps to quickly recall their primary functions.

---

130. What is the function of tassels in the corn cob?

(A) To disperse pollen grains
(B) To protect seeds
(C) To attract insects
(D) To trap pollen grains

Correct Answer: (D) To trap pollen grains

Solution:

Step 1: Understanding the Question
The question asks for the function of "tassels in the corn cob". It's important to note the specific botanical structures of a corn plant (maize).

Step 2: Detailed Explanation
There appears to be a biological inaccuracy in the question's phrasing. In a corn plant:
Tassel: This is the male inflorescence located at the top of the plant. Its primary function is to produce and disperse a large amount of pollen grains. This corresponds to option (A).
Corn Cob/Ear: This is the female inflorescence located lower on the stalk. It develops into the fruit (the corn we eat). From the tip of the cob emerge long, silky threads called **silks**.
Silks: Each silk is a stigma and style. The function of the silks is to trap the airborne pollen grains dispersed from the tassels. This corresponds to option (D).

The question incorrectly links "tassels" with the "corn cob". However, given the options, the question is likely asking about the function of the structures associated with the cob that are involved in pollination, which are the silks. The function of the silks is to trap pollen grains. The provided answer key indicates (D) is correct, reinforcing the interpretation that the question, despite its flawed terminology, refers to the function of the silks on the cob.

Step 3: Final Answer
Assuming the question mistakenly refers to the silks on the corn cob instead of the tassels, their function is to trap pollen grains. Therefore, option (D) is the intended correct answer.

Quick Tip

In competitive exams, be aware of potentially poorly phrased questions. Analyze all options and choose the one that is most biologically plausible in the context, even if the terminology is slightly incorrect. In maize, remember: Tassel (male, top) disperses pollen; Silk (female, on cob) traps pollen.

---

131. In gene gun method used to introduce alien DNA into host cells, microparticles of ___ metal are used.

(A) Tungsten or gold
(B) Silver
(C) Copper
(D) Zinc

Correct Answer: (A) Tungsten or gold

Solution:

Step 1: Understanding the Question
The question asks to identify the metals used as microparticles in the gene gun method for genetic transformation.

Step 2: Detailed Explanation
The gene gun method, also known as biolistics or microprojectile bombardment, is a direct method of gene transfer, primarily used for transforming plant cells.
Principle: The method involves coating microscopic particles of a heavy metal with the desired DNA (the "alien" DNA or transgene). These DNA-coated microparticles are then accelerated to a high velocity and fired into the target host cells or tissues. The particles penetrate the cell walls and membranes, carrying the DNA into the cell's interior, where it can be incorporated into the host genome.
Choice of Metal: The microparticles must be dense enough to achieve the necessary momentum to penetrate the cell wall, but also small enough not to cause excessive damage. Most importantly, they must be biologically inert so they don't cause a toxic reaction within the cell. **Gold (Au)** and **Tungsten (W)** fit these criteria perfectly. They are very dense and chemically non-reactive. Silver, copper, and zinc are generally too reactive or toxic to be used for this purpose.

Step 3: Final Answer
Gold and tungsten are the standard metals used for coating with DNA in the gene gun method. Therefore, option (A) is the correct answer.

> **Quick Tip**
>
> Associate "gene gun" with "golden bullets". This mnemonic helps to remember that gold is one of the primary metals used, along with tungsten, for the microprojectiles in biolistics.

---

132. Cellulose does not form blue colour with Iodine because

(A) It does not contain complex helices and hence cannot hold iodine molecules.
(B) It breakes down when iodine reacts with it.
(C) It is a disaccharide.
(D) It is a helical molecule.

Correct Answer: (A) It does not contain complex helices and hence cannot hold iodine molecules.

Solution:

Step 1: Understanding the Question
The question asks for the reason why cellulose does not give a positive result (blue-black colour) with the iodine test, unlike starch.

Step 2: Detailed Explanation
The iodine test is specific for the presence of starch. The principle behind this test lies in the structure of the polysaccharide.
Starch: Starch consists of two components, amylose and amylopectin. Amylose is an unbranched polymer of α-glucose units linked by α-1,4 glycosidic bonds. This structure causes the amylose chain to form a complex helical (coiled) shape. Iodine molecules (specifically, the I_3^- and I_5^- ions in the iodine solution) can fit inside this helix, forming a starch-iodine complex. This complex absorbs light, resulting in the characteristic blue-black colour.
Cellulose: Cellulose is a polymer of β-glucose units linked by β-1,4 glycosidic bonds. This type of linkage results in a straight, linear chain rather than a helix. The chains are arranged parallel to each other and are held by extensive hydrogen bonds, forming strong microfibrils. Because cellulose does not have the complex helical structure of amylose, it cannot trap or hold iodine molecules to form the colored complex.

Analyzing the options:
(A) This correctly states that cellulose lacks complex helices and thus cannot hold iodine.
(B) Cellulose is a very stable polymer and does not break down upon reaction with iodine.
(C) Cellulose is a polysaccharide, not a disaccharide.
(D) Cellulose is a linear, uncoiled molecule, not helical.

Step 3: Final Answer
The absence of a helical structure in cellulose prevents the formation of an iodine complex, so it does not turn blue. Thus, option (A) is correct.

Quick Tip

Remember the key structural difference: Starch (α-glucose) = Helical = Traps Iodine = Blue colour. Cellulose (β-glucose) = Linear = No trapping = No blue colour. This structural difference also explains why humans can digest starch but not cellulose.

---

133. Which of the following stages of meiosis involves division of centromere?

(A) Anaphase II
(B) Telophase
(C) Metaphase I
(D) Metaphase II

Correct Answer: (A) Anaphase II

Solution:

Step 1: Understanding the Question
The question asks to identify the specific stage in the meiotic process where the centromeres split.

Step 2: Differentiating Meiosis I and Meiosis II
Meiosis consists of two successive divisions, Meiosis I and Meiosis II.
Meiosis I (Reductional Division): The primary event in Anaphase I is the separation of **homologous chromosomes**. The sister chromatids of each chromosome remain attached at their centromeres. Thus, the centromeres do **not** divide in Meiosis I.
Meiosis II (Equational Division): This division is very similar to mitosis. In Metaphase II, individual chromosomes (each consisting of two sister chromatids) align at the equatorial plate. In **Anaphase II**, the centromere of each chromosome finally divides, allowing the **sister chromatids** to separate and move to opposite poles. These separated chromatids are now considered individual chromosomes.

Step 3: Analyzing the Stages
- **Metaphase I:** Homologous pairs align; centromeres do not divide.
- **Metaphase II:** Individual chromosomes align; centromeres do not divide yet.
- **Anaphase II:** Centromeres divide, and sister chromatids separate.
- **Telophase:** Division is nearly complete; this phase follows the separation of chromatids.

Step 4: Final Answer
The division of the centromere, which leads to the separation of sister chromatids, occurs during Anaphase II. Therefore, option (A) is correct.

> Quick Tip
>
> A key difference to remember: - **Anaphase I** separates **homologous chromosomes**. - **Anaphase II** separates **sister chromatids** (due to centromere division). This distinction is fundamental to understanding meiosis.

---

134. The reaction centre in PS II has an absorption maxima at

(A) 660 nm
(B) 780 nm
(C) 680 nm
(D) 700 nm

Correct Answer: (C) 680 nm

Solution:

Step 1: Understanding the Question
The question asks for the specific wavelength of light at which the reaction centre of Photosystem II (PS II) shows maximum absorption.

Step 2: Detailed Explanation
In photosynthesis, there are two photosystems, PS I and PS II, each with a reaction centre composed of a special pair of chlorophyll 'a' molecules. These reaction centres are named after their characteristic absorption peaks in the red region of the light spectrum.
Photosystem II (PS II): Its reaction centre is called **P680** because it absorbs light most effectively at a wavelength of **680 nm**.
Photosystem I (PS I): Its reaction centre is called **P700** because it absorbs light most effectively at a wavelength of **700 nm**.

Therefore, the absorption maximum for the reaction centre in PS II is 680 nm.

Step 3: Final Answer
The reaction centre in PS II is known as P680, indicating its absorption maxima is at 680 nm. Thus, option (C) is correct.

Quick Tip

A simple way to remember is that in the non-cyclic electron flow (Z-scheme), PS II comes before PS I. Similarly, the number 680 comes before 700. So, PS II corresponds to P680 and PS I corresponds to P700.

---

135. The historic Convention on Biological Diversity, 'The Earth Summit' was held in Rio de Janeiro in the year :

(A) 1986
(B) 2002
(C) 1985
(D) 1992

Correct Answer: (D) 1992

Solution:

Step 1: Understanding the Question
The question asks for the year in which the Earth Summit, where the Convention on Biological Diversity (CBD) was established, was held in Rio de Janeiro.

Step 2: Detailed Explanation
The United Nations Conference on Environment and Development (UNCED), popularly known as the Earth Summit, was a major international conference held in Rio de Janeiro, Brazil.

This summit took place from June 3 to June 14, 1992.
One of the key outcomes of this summit was the opening for signature of the Convention on Biological Diversity (CBD).
Therefore, the historic event occurred in 1992.

Step 3: Final Answer
Based on historical facts, the Earth Summit in Rio de Janeiro was held in 1992. Hence, option (D) is the correct answer.

Quick Tip

Memorize the years and locations of major environmental summits and protocols, such as the Earth Summit (1992, Rio), the World Summit on Sustainable Development (2002, Johannesburg), the Kyoto Protocol (1997), and the Montreal Protocol (1987).

---

136. Match List I with List II :

	List I		List II
A.	**M Phase**	**I.**	**Proteins are synthesized**
B.	**G_2 Phase**	**II.**	**Inactive phase**
C.	**Quiescent stage**	**III.**	**Interval between mitosis and initiation of DNA replication**
D.	**G_1 Phase**	**IV.**	**Equational division**

Choose the correct answer from the options given below :

(A) A-IV, B-I, C-II, D-III
(B) A-II, B-IV, C-I, D-III
(C) A-III, B-II, C-IV, D-I
(D) A-IV, B-II, C-I, D-III

Correct Answer: (A) A-IV, B-I, C-II, D-III

Solution:

Step 1: Understanding the Question
The question requires matching the phases of the cell cycle (List I) with their corresponding events or descriptions (List II).

Step 2: Detailed Matching
A. M Phase: This is the mitotic phase where the cell divides. Mitosis is known as **equational division** because the daughter cells have the same number of chromosomes as the parent cell. Thus, **A matches with IV**.
B. G_2 Phase: This is the second gap phase, occurring after DNA synthesis (S phase) and before mitosis (M phase). During this phase, the cell continues to grow, and crucial **proteins are synthesized**, such as tubulin for the mitotic spindle. Thus, **B matches with I**.

C. Quiescent stage (G_0): This is a state where cells exit the cell cycle and stop dividing. They are metabolically active but are in an **inactive phase** with respect to proliferation. Thus, **C matches with II**.
D. G_1 Phase: This is the first gap phase, which is the **interval between** the end of **mitosis** (M phase) and the **initiation of DNA replication** (S phase). Thus, **D matches with III**.

Step 3: Final Answer
The correct set of matches is: A $\rightarrow$ IV, B $\rightarrow$ I, C $\rightarrow$ II, D $\rightarrow$ III. This corresponds to option (A).

Quick Tip

Draw the cell cycle diagram (a circle with G_1, S, G_2, M phases) and label the key event of each phase. This visual aid makes matching questions much easier to solve. G_0 is an exit from the G_1 phase.

---

137. Given below are two statements :
Statement I: Gause's 'Competitive Exclusion Principle' states that two closely related species competing for the same resources cannot co-exist indefinitely and competitively inferior one will be eliminated eventually.
Statement II: In general, carnivores are more adversely affected by competition than herbivores.
In the light of the above statements, choose the correct answer from the options given below :

(A) Statement I is correct but Statement II is false.
(B) Statement I is incorrect but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (1) Statement I is correct but Statement II is false.

Solution:

Step 1: Understanding the Question
The question presents two statements related to ecological competition and asks us to evaluate their correctness.

Step 2: Detailed Explanation
Analysis of Statement I:
Statement I provides a definition of Gause's 'Competitive Exclusion Principle'. This principle posits that when two species compete for the exact same limited resources within a stable environment, one will be more efficient and will eventually outcompete and eliminate the other. This definition is accurate. A classic example is the experiment with two species of *Paramecium*, *P. aurelia* and *P. caudatum*. When grown together, *P. aurelia* outcompeted *P.

caudatum*. Thus, **Statement I is correct**.

Analysis of Statement II:
Statement II claims that carnivores are generally more adversely affected by competition than herbivores. This is a generalization that is not universally accepted and is often considered incorrect in ecological theory. Competition can be intense at all trophic levels. Herbivores often face strong competition for limited plant resources, which can be just as, if not more, intense than competition among carnivores for prey. For instance, competition for grazing land among different herbivore species can be very high. Therefore, making a blanket statement that one group is "more adversely affected" is not accurate. Thus, **Statement II is false**.

Step 3: Final Answer
Since Statement I is a correct definition of Gause's principle and Statement II is an incorrect generalization, the correct option is (1).

> **Quick Tip**
>
> Remember Gause's principle as "one niche, one species". However, also be aware of mechanisms that allow coexistence, such as resource partitioning (e.g., MacArthur's warblers feeding in different parts of the same tree), which is an exception to competitive exclusion.

---

138. How many different proteins does the ribosome consist of?

(A) 40
(B) 20
(C) 80
(D) 60

Correct Answer: (C) 80

Solution:

Step 1: Understanding the Question
The question asks for the approximate number of different proteins found in a ribosome. As it does not specify prokaryotic or eukaryotic, we consider the most common context in general biology, which is the eukaryotic ribosome.

Step 2: Detailed Explanation
Ribosomes are complex molecular machines made of ribosomal RNA (rRNA) and ribosomal proteins.
Eukaryotic Ribosome (80S): It is composed of two subunits:
- **Large Subunit (60S):** Contains 3 types of rRNA molecules and approximately 49 different proteins.

- **Small Subunit (40S):** Contains 1 type of rRNA molecule and approximately 33 different proteins.
The total number of different proteins in an 80S eukaryotic ribosome is therefore approximately $49 + 33 = 82$ proteins.

Prokaryotic Ribosome (70S): For comparison, it consists of:
- Large Subunit (50S): 34 proteins.
- Small Subunit (30S): 21 proteins.
Total proteins are approximately 55.

Looking at the options, the value '80' is a very close approximation for the number of proteins in a eukaryotic ribosome.

Step 3: Final Answer
A eukaryotic ribosome is composed of approximately 80 different proteins. Thus, option (C) is the correct answer.

Quick Tip

For questions on ribosomes, remember the Svedberg units and composition. Eukaryotes: 80S = 60S + 40S, with 80 proteins. Prokaryotes: 70S = 50S + 30S, with 55 proteins. The numbers don't add up arithmetically because Svedberg units are a measure of sedimentation rate, not mass.

---

139. Match List I with List II :

List I		List II	
A.	Oxidative decarboxylation	I.	Citrate synthase
B.	Glycolysis	II.	Pyruvate dehydrogenase
C.	Oxidative phosphorylation	III.	Electron transport system
D.	Tricarboxylic acid cycle	IV.	EMP pathway

Choose the correct answer from the options given below :

(A) A-III, B-I, C-II, D-IV
(B) A-II, B-IV, C-III, D-I
(C) A-III, B-IV, C-II, D-I
(D) A-II, B-IV, C-I, D-III

Correct Answer: (2) A-II, B-IV, C-III, D-I

Solution:

Step 1: Understanding the Question
The question requires matching metabolic processes or reactions from List I with the associated

enzyme, pathway, or system from List II.

Step 2: Detailed Explanation
Let's match each item from List I with its correct counterpart in List II.

- **A. Oxidative decarboxylation:** This is a key reaction that links glycolysis to the Krebs cycle. It involves the conversion of pyruvate to acetyl-CoA, a process catalyzed by the **Pyruvate dehydrogenase** complex. This matches with **II**.
- **B. Glycolysis:** This is the initial pathway of glucose breakdown. It is also known as the **EMP pathway**, named after its discoverers Embden, Meyerhof, and Parnas. This matches with **IV**.
- **C. Oxidative phosphorylation:** This is the process where ATP is formed as a result of the transfer of electrons from NADH or $FADH_2$ to O_2 by a series of electron carriers. This process takes place in the **Electron transport system** (ETS). This matches with **III**.
- **D. Tricarboxylic acid (TCA) cycle:** Also known as the Krebs cycle or citric acid cycle. The very first step of this cycle is the condensation of acetyl-CoA and oxaloacetate to form citrate, a reaction catalyzed by the enzyme **Citrate synthase**. This matches with **I**.

Step 3: Final Answer
Based on the matching:
A matches with II.
B matches with IV.
C matches with III.
D matches with I.
This combination corresponds to **A-II, B-IV, C-III, D-I**, which is option (2).

Quick Tip

Associate key enzymes or alternative names with major metabolic pathways. For example, Glycolysis = EMP pathway; Krebs Cycle starts with Citrate Synthase; the Link Reaction is catalyzed by Pyruvate Dehydrogenase; and ATP synthesis via ETS is Oxidative Phosphorylation.

---

140. Which of the following statements are correct about Klinefelter's Syndrome?
A. This disorder was first described by Langdon Down (1866).
B. Such an individual has overall masculine development. However, the feminine development is also expressed.
C. The affected individual is short statured.
D. Physical, psychomotor and mental development is retarded.

E. Such individuals are sterile.
Choose the correct answer from the options given below :

(A) B and E only
(B) A and E only
(C) A and B only
(D) C and D only

Correct Answer: (1) B and E only

Solution:

Step 1: Understanding the Question
The question asks to identify the correct statements describing Klinefelter's Syndrome from a given list.

Step 2: Detailed Explanation
Klinefelter's Syndrome is a genetic disorder caused by the presence of an extra X chromosome in males, resulting in the karyotype 47, XXY. Let's evaluate each statement:

- **Statement A:** This is incorrect. Langdon Down described Down's Syndrome (Trisomy 21). Klinefelter's Syndrome was described by Harry Klinefelter in 1942.

- **Statement B:** This is correct. Individuals with Klinefelter's Syndrome are phenotypically male and have overall masculine development. However, the extra X chromosome leads to the expression of some feminine characteristics, such as the development of breasts (gynaecomastia).

- **Statement C:** This is incorrect. Individuals with Klinefelter's Syndrome are often taller than average, not short statured. Short stature is a characteristic of Turner's Syndrome (45, XO).

- **Statement D:** This is incorrect. This description is more characteristic of Down's Syndrome. While some learning disabilities may be present in individuals with Klinefelter's Syndrome, severe retardation of physical, psychomotor, and mental development is not a typical feature.

- **Statement E:** This is correct. The presence of the extra X chromosome leads to underdeveloped testes (testicular atrophy), resulting in low testosterone production and sterility.

Step 3: Final Answer
Based on the analysis, only statements B and E are correct. Therefore, option (1) is the correct answer.

Quick Tip

To avoid confusion between chromosomal disorders, create a summary table. For each syndrome (e.g., Down's, Turner's, Klinefelter's), list the karyotype, key physical features, and effects on development and fertility.

---

141. Malonate inhibits the growth of pathogenic bacteria by inhibiting the activity of

(A) Lipase
(B) Dinitrogenase
(C) Succinic dehydrogenase
(D) Amylase

Correct Answer: (3) Succinic dehydrogenase

Solution:

Step 1: Understanding the Question
The question asks for the enzyme that is inhibited by malonate, leading to the inhibition of bacterial growth. This points towards a specific type of enzyme inhibition.

Step 2: Key Concepts
This question relates to the concept of **competitive enzyme inhibition**. A competitive inhibitor is a molecule that structurally resembles the enzyme's substrate. It competes with the substrate for binding to the active site of the enzyme. By binding to the active site, the inhibitor prevents the substrate from binding, thereby reducing the enzyme's activity.

Step 3: Detailed Explanation

- **Enzyme and Substrate:** The enzyme in question is **succinic dehydrogenase**. It is a key enzyme in the Krebs cycle (citric acid cycle), which is a central metabolic pathway for energy production in aerobic organisms, including many pathogenic bacteria.

- **Substrate:** The natural substrate for succinic dehydrogenase is **succinate**. The enzyme catalyzes the oxidation of succinate to fumarate.

$$\text{Succinate} \xrightarrow{\text{Succinic dehydrogenase}} \text{Fumarate}$$

- **Inhibitor: Malonate** (or malonic acid) has a chemical structure very similar to succinate.

Due to this structural similarity, malonate acts as a competitive inhibitor of succinic dehydrogenase. It binds to the active site of the enzyme but cannot be acted upon. This blocks the active site and prevents succinate from binding. By inhibiting this crucial step in the Krebs cycle, malonate disrupts cellular respiration and ATP production, which in turn inhibits the growth and proliferation of the bacteria.

Step 4: Final Answer
Malonate is a classic competitive inhibitor of the enzyme succinic dehydrogenase. Therefore, option (3) is the correct answer.

> Quick Tip
>
> Remember the classic example of competitive inhibition: succinic dehydrogenase is inhibited by its structural analogue, malonate. This is a frequently tested concept in exams.

---

142. Which of the following combinations is required for chemiosmosis?

(A) proton pump, electron gradient, ATP synthase
(B) proton pump, electron gradient, NADP synthase
(C) membrane, proton pump, proton gradient, ATP synthase
(D) membrane, proton pump, proton gradient, NADP synthase

Correct Answer: (3) membrane, proton pump, proton gradient, ATP synthase

Solution:

Step 1: Understanding the Question
The question asks to identify the essential components required for the process of chemiosmosis.

Step 2: Key Concepts
Chemiosmosis is the mechanism by which ATP is produced during cellular respiration and photosynthesis. It involves the movement of ions (specifically protons or H^+) across a selectively permeable membrane, down their electrochemical gradient. This process is described by Peter Mitchell's chemiosmotic theory.

Step 3: Detailed Explanation
The key components for chemiosmosis are:
1. **A membrane:** A semipermeable membrane (like the inner mitochondrial membrane or the thylakoid membrane) is essential to establish and maintain a concentration gradient of protons.
2. **A proton pump:** This is a mechanism to actively transport protons across the membrane, from a region of low concentration to a region of high concentration. This is typically achieved by the electron transport chain (ETC), which uses the energy from electrons to pump H^+.
3. **A proton gradient:** The pumping of protons creates a high concentration of H^+ on one side of the membrane, resulting in a proton motive force (an electrochemical gradient).
4. **ATP synthase:** This is an enzyme complex embedded in the membrane that provides a channel for protons to flow back down their concentration gradient. The energy released from

this flow is used by ATP synthase to synthesize ATP from ADP and inorganic phosphate (Pi).

Analyzing the options:
(1) and (2) are incomplete as they miss the essential membrane component required to maintain the gradient.
(3) includes all four necessary components: the membrane, the pump, the resulting gradient, and the enzyme (ATP synthase) that utilizes the gradient.
(4) is incorrect because NADP synthase is not directly involved in chemiosmosis for ATP synthesis; instead, $NADP^+$ reductase is involved in the final step of the light-dependent reactions of photosynthesis to produce NADPH.

Step 4: Final Answer
The correct combination of components required for chemiosmosis is membrane, proton pump, proton gradient, and ATP synthase. Thus, option (3) is the correct answer.

> **Quick Tip**
>
> Think of chemiosmosis like a dam. The membrane is the dam wall, the proton pump is the machinery that fills the reservoir (creating the proton gradient), and the ATP synthase is the turbine through which water flows to generate electricity (ATP).

---

143. Match List I with List II:

List I		List II	
A.	Cohesion	I.	More attraction in liquid phase
B.	Adhesion	II.	Mutual attraction among water molecules
C.	Surface tension	III.	Water loss in liquid phase
D.	Guttation	IV.	Attraction towards polar surfaces

Choose the correct answer from the options given below :

(A) A-III, B-I, C-IV, D-II
(B) A-II, B-I, C-IV, D-III
(C) A-II, B-IV, C-I, D-III
(D) A-IV, B-III, C-II, D-I

Correct Answer: (3) A-II, B-IV, C-I, D-III

Solution:

Step 1: Understanding the Question
The question requires matching terms related to the properties of water and plant water relations (List I) with their correct definitions or descriptions (List II).

Step 2: Detailed Explanation
Let's match each term in List I with its correct definition in List II.

- **A. Cohesion:** This is the property of like molecules sticking to each other due to mutual attraction. For water, it's the attraction among water molecules. This matches with **II. Mutual attraction among water molecules**.

- **B. Adhesion:** This is the property of different molecules or surfaces clinging to one another. In plants, it refers to the attraction of water molecules to the polar surfaces of xylem elements. This matches with **IV. Attraction towards polar surfaces**.

- **C. Surface tension:** This property is a direct result of cohesion. Water molecules at the surface are more strongly attracted to other water molecules in the liquid phase than to the molecules in the air above. This matches with **I. More attraction in liquid phase**.

- **D. Guttation:** This is the process of exudation of water droplets (xylem sap) from the tips or margins of leaves, typically occurring at night when transpiration is low and root pressure is high. It is a form of water loss in the liquid phase. This matches with **III. Water loss in liquid phase**.

Step 3: Final Answer
Based on the matching:
A matches with II.
B matches with IV.
C matches with I.
D matches with III.
This combination corresponds to **A-II, B-IV, C-I, D-III**, which is option (3).

> Quick Tip
>
> Remember the 'Co-' in Cohesion means 'together' (like molecules together), and 'Ad-' in Adhesion means 'to' (sticking 'to' a different surface). Guttation is often confused with dew, but guttation is water coming from *inside* the plant, whereas dew is condensation from the atmosphere.

---

144. Match List I with List II :

List I		**List II**	
A.	Iron	I.	Synthesis of auxin
B.	Zinc	II.	Component of nitrate reductase
C.	Boron	III.	Activator of catalase
D.	Molybdenum	IV.	Cell elongation and differentiation

Choose the correct answer from the options given below :

(A) A-III, B-I, C-IV, D-II
(B) A-II, B-IV, C-I, D-III
(C) A-III, B-II, C-I, D-IV
(D) A-II, B-III, C-IV, D-I

Correct Answer: (1) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question
The question requires matching the micronutrients in List I with their specific functions in plants from List II.

Step 2: Detailed Explanation
Let's match each micronutrient with its correct function.

- **A. Iron (Fe):** Iron is a crucial component of proteins involved in redox reactions, such as cytochromes in the electron transport chain. It is also essential for the formation of chlorophyll and is a key part of the enzyme catalase, being necessary for its activation and function. This matches with **III. Activator of catalase**.

- **B. Zinc (Zn):** Zinc is required for the activity of various enzymes, especially carboxylases. It is also critically needed for the synthesis of auxin, a major plant growth hormone. This matches with **I. Synthesis of auxin**.

- **C. Boron (B):** Boron is required for the uptake and utilization of Ca^{2+}, membrane functioning, pollen germination, cell elongation, and cell differentiation. This matches with **IV. Cell elongation and differentiation**.

- **D. Molybdenum (Mo):** Molybdenum is a component of several enzymes, including nitrogenase and nitrate reductase, both of which are critical for nitrogen metabolism in plants. This matches with **II. Component of nitrate reductase**.

Step 3: Final Answer
Based on the matching:
A matches with III.
B matches with I.
C matches with IV.
D matches with II.
This combination corresponds to **A-III, B-I, C-IV, D-II**, which is option (1).

Quick Tip

Create flashcards for essential micronutrients and their key functions. Mnemonics can be helpful, for example: "ZinC for AuXin" and "Molybdenum for Nitrogen metabolism (Nitrate Reductase, Nitrogenase)".

---

145. Main steps in the formation of Recombinant DNA are given below. Arrange these steps in a correct sequence.

A. Insertion of recombinant DNA into the host cell.
B. Cutting of DNA at specific location by restriction enzyme.
C. Isolation of desired DNA fragment.
D. Amplification of gene of interest using PCR.
Choose the correct answer from the options given below :

(A) C, B, D, A
(B) B, D, A, C
(C) B, C, D, A
(D) C, A, B, D

Correct Answer: (1) C, B, D, A

Solution:

Step 1: Understanding the Question
The question asks to arrange the given steps of creating recombinant DNA in the correct chronological order.

Step 2: Detailed Explanation
Let's break down the logical flow of recombinant DNA technology:
1. **Isolation of Genetic Material (DNA):** The very first step is to isolate the desired DNA (the gene of interest) from the source organism. This corresponds to step **C. Isolation of desired DNA fragment.**
2. **Cutting the DNA:** Once the DNA is isolated, both the gene of interest and the vector DNA (e.g., a plasmid) must be cut with the same restriction enzyme to create complementary "sticky ends". This corresponds to step **B. Cutting of DNA at specific location by restriction enzyme.**
3. **Amplification of Gene of Interest:** To obtain a sufficient quantity of the desired gene for ligation, it is amplified using the Polymerase Chain Reaction (PCR). This step ensures there are many copies of the gene to be inserted into the vectors. This corresponds to step **D. Amplification of gene of interest using PCR.**
4. **Ligation:** The amplified gene of interest is then joined (ligated) with the cut vector DNA using the enzyme DNA ligase. This forms the recombinant DNA molecule. (This step is implicit between D and A).
5. **Transformation/Insertion:** The final step in this sequence is to introduce the recombinant DNA into a suitable host cell (like E. coli) where it can replicate. This process is called trans-

formation. This corresponds to step **A. Insertion of recombinant DNA into the host cell.**

Step 3: Final Answer
The correct sequence of the given steps is C $\rightarrow$ B $\rightarrow$ D $\rightarrow$ A. This matches option (1).

> Quick Tip
>
> Remember the acronym "I-C-A-L-I": **I**solation, **C**utting, **A**mplification, **L**igation, **I**nsertion. This covers the main workflow for creating a recombinant organism. (Ligation is not an option here but happens between D and A).

---

146. Match List I with List II :

List I (Interaction)		List II (Species A and B)	
A.	Mutualism	I.	+(A), O(B)
B.	Commensalism	II.	-(A), O(B)
C.	Amensalism	III.	+(A), -(B)
D.	Parasitism	IV.	+(A), +(B)

Choose the correct answer from the options given below :

(A) A-IV, B-III, C-I, D-II
(B) A-III, B-I, C-IV, D-II
(C) A-IV, B-II, C-I, D-III
(D) A-IV, B-I, C-II, D-III

Correct Answer: (4) A-IV, B-I, C-II, D-III

Solution:

Step 1: Understanding the Question
The question requires matching different types of ecological interactions (List I) with their symbolic representation (List II), where '+' denotes benefit, '-' denotes harm, and 'O' denotes no effect.

Step 2: Detailed Explanation
Let's define each interaction and match it to its symbol.

- **A. Mutualism:** An interaction where both species (A and B) benefit from the relationship. This is represented as (+, +). This matches with **IV. +(A), +(B)**.

- **B. Commensalism:** An interaction where one species (A) benefits, and the other species (B) is neither harmed nor benefited (unaffected). This is represented as (+, O). This matches with **I. +(A), O(B)**.

- **C. Amensalism:** An interaction where one species (A) is harmed, and the other species (B) is unaffected. This is represented as (-, O). This matches with **II. -(A), O(B)**.

- **D. Parasitism:** An interaction where one species (the parasite, A) benefits at the expense of the other species (the host, B), which is harmed. This is represented as (+, -). This matches with **III. +(A), -(B)**. (Note: The question paper OCR might have a typo for option III, but based on standard definitions, this is the correct match).

Step 3: Final Answer
Based on the matching:
A matches with IV.
B matches with I.
C matches with II.
D matches with III.
This combination corresponds to **A-IV, B-I, C-II, D-III**, which is option (4).

> **Quick Tip**
>
> Create a table for all population interactions (Mutualism, Commensalism, Amensalism, Parasitism, Predation, Competition) and their (+, -, O) notation. This makes it easy to memorize and quickly answer matching questions.

---

147. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R:
Assertion A: A flower is defined as modified shoot wherein the shoot apical meristem changes to floral meristem.
Reason R: Internode of the shoot gets condensed to produce different floral appendages laterally at successive nodes instead of leaves.
In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (C) Both A and R are true and R is the correct explanation of A.

Solution:

Step 1: Understanding the Statements
The question presents an Assertion (A) defining a flower and a Reason (R) explaining the

structural modification. We need to evaluate both statements and their relationship.

Step 2: Detailed Explanation
Assertion A: "A flower is defined as modified shoot wherein the shoot apical meristem changes to floral meristem." This is the correct and standard botanical definition of a flower. The transition from a vegetative phase to a reproductive phase involves the transformation of the shoot apical meristem into a floral meristem, which has determinate growth. So, **Assertion A is true**.
Reason R: "Internode of the shoot gets condensed to produce different floral appendages laterally at successive nodes instead of leaves." This statement accurately describes the modification process. The floral axis, or thalamus, is essentially a shoot with highly condensed internodes. The floral whorls (sepals, petals, stamens, and carpels) are homologous to leaves and are arranged at these closely packed nodes. So, **Reason R is also true**.

Connecting A and R:
The Reason (condensation of internodes and production of floral appendages instead of leaves) explains exactly why a flower is considered a modified shoot (the Assertion). It details the structural changes that occur when a shoot apex becomes a floral apex. Therefore, **Reason R is the correct explanation of Assertion A**.

Step 3: Final Answer
Both A and R are true, and R correctly explains A. Therefore, option (C) is the correct answer.

> Quick Tip
>
> Evidence for the "flower as a modified shoot" theory includes that floral parts sometimes revert to leaf-like structures (phyllody) and that the thalamus is essentially a stem with nodes and condensed internodes.

---

148. Which one of the following statements is NOT correct?

(A) Water hyacinth grows abundantly in eutrophic water bodies and leads to an imbalance in the ecosystem dynamics of the water body.
(B) The amount of some toxic substances of industrial waste water increases in the organisms at successive trophic levels.
(C) The micro-organisms involved in biodegradation of organic matter in a sewage polluted water body consume a lot of oxygen causing the death of aquatic organisms.
(D) Algal blooms caused by excess of organic matter in water improve water quality and promote fisheries.

Correct Answer: (4) Algal blooms caused by excess of organic matter in water improve water quality and promote fisheries.

Solution:

Step 1: Understanding the Question
The question asks to identify the incorrect statement among the given options related to water pollution and its effects on aquatic ecosystems.

Step 2: Detailed Explanation
Let's analyze each statement:
(1) **Water hyacinth in eutrophic water bodies:** This statement is correct. Eutrophic water bodies are rich in nutrients, which promotes the excessive growth of aquatic plants like water hyacinth. This leads to an imbalance in the ecosystem.
(2) **Biomagnification:** This statement describes biomagnification (or bioaccumulation), where the concentration of toxic substances (like heavy metals or pesticides) increases at successive trophic levels in a food chain. This is a correct phenomenon.
(3) **Biodegradation and Oxygen Depletion:** This statement is correct. When sewage with a high amount of organic matter is discharged into a water body, decomposer microorganisms break it down. This process consumes a large amount of dissolved oxygen, leading to a sharp drop in oxygen levels (measured as high Biological Oxygen Demand or BOD). The lack of oxygen can cause the death of fish and other aquatic organisms.
(4) **Algal Blooms:** This statement is incorrect. Algal blooms are caused by an excess of nutrients (like nitrates and phosphates), not primarily organic matter, in the water. These blooms drastically deteriorate water quality. When the algae die, they are decomposed by bacteria, which consumes a large amount of dissolved oxygen, leading to hypoxia or anoxia and the death of fish. Therefore, algal blooms harm fisheries, not promote them, and degrade water quality.

Step 3: Final Answer
The statement that algal blooms improve water quality and promote fisheries is factually incorrect. Hence, option (4) is the correct answer.

Quick Tip

Remember the key consequences of eutrophication: excessive plant/algal growth, increased BOD, oxygen depletion, and death of aquatic animals. Algal blooms are a clear indicator of poor water quality.

---

149. Identify the correct statements :
A. Lenticels are the lens-shaped openings permitting the exchange of gases.
B. Bark formed early in the season is called hard bark.
C. Bark is a technical term that refers to all tissues exterior to vascular cambium.
D. Bark refers to periderm and secondary phloem.
E. Phellogen is single-layered in thickness.
Choose the correct answer from the options given below :

(A) A, B and D only
(B) B and C only

(C) B, C and E only
(D) A and D only

Correct Answer: (4) A and D only

Solution:

Step 1: Understanding the Question
The question asks to identify which of the five given statements about plant anatomy (specifically bark and related structures) are correct.

Step 2: Detailed Explanation
Let's evaluate each statement:

- **Statement A:** Lenticels are indeed lens-shaped pores on the bark of woody plants that allow for the exchange of gases between the internal tissues and the atmosphere. **This statement is correct.**

- **Statement B:** Bark formed early in the season (spring wood) is known as 'soft bark', while bark formed later in the season (autumn wood) is called 'hard bark'. Therefore, this statement is incorrect.

- **Statement C:** While 'bark' is a non-technical term that broadly refers to all tissues outside the vascular cambium, this definition can be considered less precise than statement D. In the context of multiple-choice questions where a more specific correct option exists, the general one might be excluded.

- **Statement D:** Anatomically, bark is composed of the periderm (cork, cork cambium, and secondary cortex) and the secondary phloem. This is a precise and correct definition of bark. **This statement is correct.**

- **Statement E:** Phellogen (cork cambium) is a meristematic tissue. It is generally described as being a single layer of cells, but it can be a few layers thick in some species. Given the definite correctness of A and D, and the ambiguity or potential inaccuracy of B, C, and E in a strict sense, we must choose the best option.

Step 3: Final Answer
Statements A and D are unequivocally correct descriptions used in botany. Statement B is incorrect. Statements C and E are debatable or less precise. Comparing the options, the combination of the most accurate statements is A and D.
Therefore, option (4) is the correct choice.

Quick Tip

In plant anatomy questions, pay close attention to precise definitions. 'Bark' has both a general and a specific anatomical definition. The specific one (Periderm + Secondary Phloem) is often preferred in exams.

---

150. Given below are two statements: One is labelled as Assertion A and the other is labelled as Reason R :

Assertion A: In gymnosperms the pollen grains are released from the microsporangium and carried by air currents.

Reason R: Air currents carry the pollen grains to the mouth of the archegonia where the male gametes are discharged and pollen tube is not formed.

In the light of the above statements, choose the correct answer from the options given below :

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (1) A is true but R is false.

Solution:

Step 1: Understanding the Question

The question asks to evaluate an Assertion (A) and a Reason (R) related to pollination and fertilization in gymnosperms.

Step 2: Detailed Explanation

Analysis of Assertion A:

Assertion A states that gymnosperm pollen grains are released from the microsporangium and carried by air currents. This describes anemophily, or wind pollination, which is the characteristic mode of pollination in most gymnosperms (like conifers). So, **Assertion A is true**.

Analysis of Reason R:

Reason R describes the events after pollination. It states that air currents carry the pollen grains to the mouth of the archegonia, and then the male gametes are discharged without the formation of a pollen tube. This is incorrect. In gymnosperms, the pollen grain lands on the micropyle of the ovule (not directly on the archegonium). It then germinates to form a **pollen tube**, which grows through the nucellus and delivers the non-motile male gametes to the vicinity of the egg cell within the archegonium. The statement "pollen tube is not formed" is a critical error. So, **Reason R is false**.

Step 3: Final Answer

Since the Assertion is true and the Reason is false, the correct option is (1).

Quick Tip

A key feature of seed plants (gymnosperms and angiosperms) is siphonogamy - the formation of a pollen tube to deliver male gametes. This adaptation eliminated the need for water for fertilization, which was required in bryophytes and pteridophytes.

Zoology

151. Match List I with List II.

List I		**List II**	
A.	Heroin	I.	Effect on cardiovascular system
B.	Marijuana	II.	Slow down body function
C.	Cocaine	III.	Painkiller
D.	Morphine	IV.	Interfere with transport of dopamine

Choose the correct answer from the options given below:

(A) A-IV, B-III, C-II, D-I
(B) A-III, B-IV, C-I, D-II
(C) A-II, B-I, C-IV, D-III
(D) A-I, B-II, C-III, D-IV

Correct Answer: (3) A-II, B-I, C-IV, D-III

Solution:

Step 1: Understanding the Question
This question requires matching drugs from List I with their primary mode of action or effect from List II.

Step 2: Detailed Explanation
Let's match each drug with its correct description:

- **A. Heroin:** Heroin (diacetylmorphine) is a powerful opioid that acts as a central nervous system depressant. It slows down body functions like breathing and heart rate. This matches with **II. Slow down body function**.

- **B. Marijuana:** The active components in marijuana are cannabinoids, which interact with cannabinoid receptors in the brain. They have a range of effects, including known impacts on the cardiovascular system, such as increased heart rate. This matches with **I. Effect on cardiovascular system**.

- **C. Cocaine:** Cocaine is a potent central nervous system stimulant. It primarily works by blocking the reuptake of neurotransmitters like dopamine, serotonin, and norepinephrine, leading to increased concentrations in the synapse. This matches with **IV. Interfere with transport of dopamine**.

- **D. Morphine:** Morphine is a classic opioid analgesic (painkiller) derived from the opium poppy. It is highly effective in relieving severe pain. This matches with **III. Painkiller**.

Step 3: Final Answer
Based on the matching:
A matches with II.
B matches with I.
C matches with IV.
D matches with III.
This combination corresponds to **A-II, B-I, C-IV, D-III**, which is option (3).

Quick Tip

Categorize drugs into major groups: Depressants (e.g., opioids like heroin, morphine), Stimulants (e.g., cocaine, amphetamines), and Hallucinogens (e.g., LSD, cannabinoids like marijuana). This helps in quickly identifying their general effects.

---

152. Given below are two statements:
Statement I: RNA mutates at a faster rate.
Statement II: Viruses having RNA genome and shorter life span mutate and evolve faster.
In the light of the above statements, choose the correct answer from the options given below:

(A) Statement I is true but Statement II is false.
(B) Statement I false but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (3) Both Statement I and Statement II are true.

Solution:

Step 1: Understanding the Question:
The question presents two statements related to the mutation rate of RNA and its consequence for RNA viruses, and asks for an evaluation of their correctness.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** RNA is chemically less stable than DNA (due to the 2'-OH group in ribose). More importantly, the enzymes that replicate RNA (RNA-dependent RNA polymerases and reverse transcriptases) typically lack the proofreading ability that DNA polymerases have. This lack of proofreading means that errors made during replication are not corrected, leading to a much higher mutation rate in RNA genomes compared to DNA genomes. Thus, Statement I is true.

- **Analysis of Statement II:** This statement builds on the first. Viruses, by their nature, have very short generation times (life spans). When this is combined with an RNA genome that has a high mutation rate (as explained in Statement I), the result is that these viruses accumulate mutations rapidly. This high rate of mutation provides a large pool of genetic variation, allowing them to evolve very quickly. This rapid evolution enables them to adapt to new hosts or evade the host's immune system, as seen in influenza virus and HIV. Thus, Statement II is also true.

Step 3: Final Answer:
Both statements are correct and factually linked. Statement I provides the molecular reason, and Statement II describes the evolutionary consequence for a specific group of organisms (RNA viruses). Therefore, option (3) is the correct answer.

> Quick Tip
>
> The high mutation rate of RNA is a key concept in virology and evolution. It explains why we need a new flu vaccine every year and why developing drugs against viruses like HIV is so challenging. Remember: RNA is unstable and its replication is error-prone.

---

153. Given below are two statements:
Statement I: Vas deferens receives a duct from seminal vesicle and opens into urethra as the ejaculatory duct.
Statement II: The cavity of the cervix is called cervical canal which along with vagina forms birth canal.
In the light of the above statements, choose the correct answer from the options given below:

(A) Statement I is correct but Statement II is false.
(B) Statement I incorrect but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (3) Both Statement I and Statement II are true.

Solution:

Step 1: Understanding the Question:
The question asks us to evaluate two statements related to the male and female reproductive systems.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** This statement describes the path of sperm in the male reproductive system. The vas deferens is the tube that carries sperm from the epididymis. Before it reaches the prostate gland, it joins with the duct from the seminal vesicle. This combined duct is called the ejaculatory duct. The ejaculatory duct then passes through the prostate gland and empties into the prostatic urethra. This description is anatomically correct.

- **Analysis of Statement II:** This statement describes the birth canal in the female reproductive system. The cervix is the lower, narrow part of the uterus. Its internal cavity is the cervical canal. During childbirth (parturition), the fetus passes from the uterus, through the dilated cervical canal, and then through the vagina to the outside. The combination of the cervical canal and the vagina constitutes the birth canal. This description is also anatomically correct.

Step 3: Final Answer:
Both Statement I and Statement II are factually correct. Therefore, option (3) is the correct answer.

Quick Tip

Trace the pathways in reproductive systems carefully. For males: Seminiferous tubules $\rightarrow$ Rete testis $\rightarrow$ Vasa efferentia $\rightarrow$ Epididymis $\rightarrow$ Vas deferens $\rightarrow$ Ejaculatory duct $\rightarrow$ Urethra. For females during birth: Uterus $\rightarrow$ Cervical canal $\rightarrow$ Vagina.

---

154. Which of the following functions is carried out by cytoskeleton in a cell?

(A) Motility
(B) Transportation
(C) Nuclear division
(D) Protein synthesis

Correct Answer: (1) Motility

Solution:

Step 1: Understanding the Question:
The question asks to identify a function performed by the cytoskeleton of a cell from the given options.

Step 2: Detailed Explanation:
The cytoskeleton is an elaborate network of protein filaments present in the cytoplasm. It performs several key functions:

- **Mechanical Support:** It helps in maintaining the shape and structure of the cell.
- **Motility:** It is fundamentally involved in cell movement. This includes the movement of the entire cell (e.g., amoeboid movement) and the movement of cellular appendages like cilia and flagella, which are made of microtubules.
- **Intracellular Transport:** It acts as a track system for motor proteins to move organelles and vesicles within the cell. This is a form of transportation.
- **Cell Division:** During both mitosis and meiosis (nuclear divisions), microtubules of the cytoskeleton assemble to form the spindle apparatus, which is responsible for separating chromosomes.

Let's evaluate the options:
(1) **Motility:** This is a direct and major function of the cytoskeleton.
(2) **Transportation:** Refers to intracellular transport, which is also a function.
(3) **Nuclear division:** Formation of the mitotic spindle is a critical role.
(4) **Protein synthesis:** This function is carried out by ribosomes, not the cytoskeleton.

Step 3: Final Answer:
Based on the options, motility is a primary and well-established function of the cytoskeleton. Thus, option (1) is the most appropriate answer.

> **Quick Tip**
>
> Remember the three main components of the cytoskeleton and their primary roles: Microfilaments (actin) for muscle contraction and cell shape, Microtubules for organelle movement and spindle formation, and Intermediate Filaments for mechanical strength. All contribute to the cell's structure and movement (motility).

---

155. Which of the following statements are correct regarding female reproductive cycle?
A. In non-primate mammals cyclical changes during reproduction are called oestrus cycle.

B. First menstrual cycle begins at puberty and is called menopause.
C. Lack of menstruation may be indicative of pregnancy.
D. Cyclic menstruation extends between menarche and menopause.
Choose the most appropriate answer from the options given below:

(A) A, B and C only
(B) A, C and D only
(C) A and D only
(D) A and B only

Correct Answer: (2) A, C and D only

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statements about the female reproductive cycle from a given list.

Step 2: Detailed Explanation:
Let's evaluate each statement:

- **A. In non-primate mammals cyclical changes during reproduction are called oestrus cycle.** This is correct. Mammals like cows, sheep, dogs, etc., exhibit an oestrus cycle, characterized by a period of "heat" (oestrus) when the female is receptive to mating. Primates (like humans, apes) have a menstrual cycle.

- **B. First menstrual cycle begins at puberty and is called menopause.** This is incorrect. The first menstrual cycle is called **menarche**. **Menopause** is the cessation (stopping) of the menstrual cycle, which occurs around the age of 50.

- **C. Lack of menstruation may be indicative of pregnancy.** This is correct. The absence of menstruation (amenorrhea) is one of the earliest and most reliable signs of pregnancy, although it can also be caused by other factors like stress, poor health, etc.

- **D. Cyclic menstruation extends between menarche and menopause.** This is correct. The reproductive phase in a human female's life is marked by the menstrual cycle, which starts at menarche (puberty) and ends at menopause.

Therefore, statements A, C, and D are correct.

Step 3: Final Answer:
The combination of correct statements is A, C, and D, which corresponds to option (2).

Quick Tip

Remember the key terms for the human female reproductive life stages: Menarche = The beginning. Menstruation = The monthly cycle. Menopause = The end. Distinguish this from the oestrus cycle in non-primates.

---

156. Match List I with List II.

List I		List II	
A.	Vasectomy	I.	Oral method
B.	Coitus interruptus	II.	Barrier method
C.	Cervical caps	III.	Surgical method
D.	Saheli	IV.	Natural method

Choose the correct answer from the options given below:

(A) A-II, B-III, C-I, D-IV
(B) A-IV, B-II, C-I, D-III
(C) A-III, B-I, C-IV, D-II
(D) A-III, B-IV, C-II, D-I

Correct Answer: (4) A-III, B-IV, C-II, D-I

Solution:

Step 1: Understanding the Question
The question requires matching different contraceptive methods from List I with their corresponding category from List II.

Step 2: Detailed Explanation
Let's categorize each contraceptive method:

- **A. Vasectomy:** This is a permanent method of contraception for males where the vas deferens is cut and tied to prevent sperm from entering the ejaculate. This is a **Surgical method** (sterilization). This matches with **III**.

- **B. Coitus interruptus:** Also known as the withdrawal method, this involves withdrawing the penis from the vagina before ejaculation. It is classified as a **Natural method** of contraception. This matches with **IV**.

- **C. Cervical caps:** These are devices made of rubber that are inserted into the vagina to cover the cervix, preventing sperm from entering the uterus. They are a type of **Barrier method**. This matches with **II**.

- **D. Saheli:** This is a non-steroidal contraceptive pill taken once a week. It is a form of hormonal contraception administered orally. This is an **Oral method**. This matches with

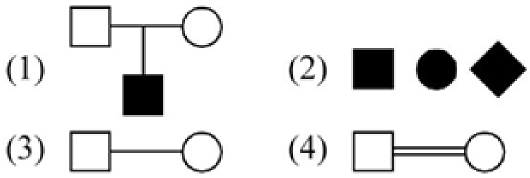
I.

Step 3: Final Answer
Based on the matching:
A matches with III.
B matches with IV.
C matches with II.
D matches with I.
This combination corresponds to **A-III, B-IV, C-II, D-I**, which is option (4).

> **Quick Tip**
>
> Organize contraceptive methods into categories: Natural (rhythm, withdrawal), Barrier (condoms, diaphragms, cervical caps), IUDs (Intra Uterine Devices), Oral (pills), and Surgical/Terminal (vasectomy, tubectomy). This structure makes matching questions easier.

---

157. Which one of the following symbols represents mating between relatives in human pedigree analysis?



(A) (1)
(B) (2)
(C) (3)
(D) (4)

Correct Answer: (2)

Solution:

Step 1: Understanding the Question
The question asks to identify the standard symbol used in human pedigree charts to represent a consanguineous mating, which is mating between relatives.

Step 2: Detailed Explanation
Let's analyze the standard symbols in pedigree analysis as represented by the options:

- **Symbol (1):** A square (male) and a circle (female) connected by a single horizontal line represents a normal mating between unrelated individuals.

- **Symbol (2):** A square and a circle connected by a **double horizontal line** is the universally accepted symbol for a consanguineous mating (mating between relatives).

- **Symbol (3):** This shows parents with an affected offspring of unspecified sex (diamond symbol). It represents a family unit, not a specific type of mating.

- **Symbol (4):** This is a more extensive pedigree chart showing a set of parents and their five offspring. It depicts a family, not the specific symbol for the type of mating.

Step 3: Final Answer
Based on established conventions in genetics and pedigree analysis, the symbol for mating between relatives is a double line connecting the male and female symbols. Therefore, the correct symbol is shown in option (2).

Quick Tip

Memorize the basic symbols for pedigree analysis: square for male, circle for female, single line for mating, double line for consanguineous mating, shaded symbol for affected individual, and a diamond for sex unspecified.

---

158. Which of the following are NOT considered as the part of endomembrane system?
A. Mitochondria
B. Endoplasmic Reticulum
C. Chloroplasts
D. Golgi complex
E. Peroxisomes
Choose the most appropriate answer from the options given below:

(A) A and D only
(B) A, D and E only
(C) B and D only
(D) A, C and E only

Correct Answer: (4) A, C and E only

Solution:

Step 1: Understanding the Question
The question asks to identify which of the listed organelles are not part of the endomembrane system.

Step 2: Detailed Explanation
The **endomembrane system** is a group of membranes and organelles in eukaryotic cells that work together to modify, package, and transport lipids and proteins. The components of this system are the nuclear envelope, the endoplasmic reticulum (ER), the Golgi apparatus (Golgi complex), lysosomes, vacuoles, and the cell membrane. They are considered a single functional unit either through direct physical continuity or by the transfer of membrane segments as vesicles.

Let's evaluate the listed organelles:

- **A. Mitochondria:** Mitochondria are involved in cellular respiration and ATP production. They are not part of the endomembrane system as their functions are distinct and not coordinated with the ER-Golgi pathway.

- **B. Endoplasmic Reticulum:** The ER is a central component of the endomembrane system, involved in protein and lipid synthesis.

- **C. Chloroplasts:** Chloroplasts are involved in photosynthesis. Like mitochondria, they are semi-autonomous organelles and are not part of the endomembrane system.

- **D. Golgi complex:** The Golgi complex is a key part of the endomembrane system, responsible for modifying, sorting, and packaging proteins and lipids for secretion or delivery to other organelles.

- **E. Peroxisomes:** Peroxisomes are small organelles that contain enzymes for metabolic processes, such as breaking down fatty acids and detoxifying harmful substances. Their functions are not coordinated with the endomembrane system.

Therefore, Mitochondria (A), Chloroplasts (C), and Peroxisomes (E) are not part of the endomembrane system.

Step 3: Final Answer
The organelles not considered part of the endomembrane system are A, C, and E. This corresponds to option (4).

Quick Tip

Remember the core components of the endomembrane system using the acronym **GERL**: **G**olgi, **E**ndoplasmic **R**eticulum, **L**ysosomes. Also include the nuclear envelope and vacuoles. Mitochondria, Chloroplasts, and Peroxisomes are the key organelles to remember as being outside this system.

---

159. Match List I with List II with respect to human eye.

List I		List II	
A.	**Fovea**	**I.**	**Visible coloured portion of eye that regulates diameter of pupil.**
B.	**Iris**	**II.**	**External layer of eye formed of dense connective tissue.**
C.	**Blind spot**	**III.**	**Point of greatest visual acuity or resolution.**
D.	**Sclera**	**IV.**	**Point where optic nerve leaves the eyeball and photoreceptor cells are absent.**

Choose the correct answer from the options given below:

(A) A-I, B-IV, C-III, D-II
(B) A-II, B-I, C-III, D-IV
(C) A-III, B-I, C-IV, D-II
(D) A-IV, B-III, C-II, D-I

Correct Answer: (3) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question:
The question requires matching different parts of the human eye (List I) with their correct descriptions or functions (List II).

Step 2: Detailed Explanation:
Let's match each part in List I with its description in List II:

- **A. Fovea:** The fovea is a small depression in the retina's macula lutea where cones are most densely packed. This concentration of cones allows for the sharpest, most detailed vision. Therefore, it is the point of greatest visual acuity or resolution. So, **A matches with III**.

- **B. Iris:** The iris is the pigmented, muscular part of the eye that gives it its color. It surrounds the pupil and controls its size by contracting or relaxing, thus regulating the amount of light entering the eye. So, **B matches with I**.

- **C. Blind spot:** The blind spot (optic disc) is the area on the retina where the ganglion cell axons exit the eye to form the optic nerve. This area lacks photoreceptor cells (rods and cones), so any image focused here cannot be seen. So, **C matches with IV**.

- **D. Sclera:** The sclera is the tough, white, fibrous outer layer of the eyeball. It is made of dense connective tissue and provides structural support and protection to the eye. So, **D matches with II**.

The correct set of matches is A-III, B-I, C-IV, D-II.

Step 3: Final Answer:
This combination corresponds to option (3).

> **Quick Tip**
>
> Create a mental map or a simple diagram of the eye to remember the location and function of each part. Fovea = Focus/Acuity. Iris = Color/Pupil control. Blind spot = Optic nerve exit/No receptors. Sclera = White/Outer protective layer.

---

160. Match List I with List II.

List I		List II	
A.	P-wave	I.	Beginning of systole
B.	Q-wave	II.	Repolarisation of ventricles
C.	QRS complex	III.	Depolarisation of atria
D.	T-wave	IV.	Depolarisation of ventricles

Choose the correct answer from the options given below:

(A) A-II, B-IV, C-I, D-III
(B) A-I, B-II, C-III, D-IV
(C) A-III, B-I, C-IV, D-II
(D) A-IV, B-III, C-II, D-I

Correct Answer: (3) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question
The question asks to match the components of a standard electrocardiogram (ECG) waveform (List I) with the cardiac event they represent (List II).

Step 2: Detailed Explanation
Let's analyze each component of the ECG:

- **A. P-wave:** This represents the electrical excitation (or depolarisation) of the atria, which leads to the contraction of both atria. This matches with **III. Depolarisation of atria**.

- **B. Q-wave:** The Q-wave is the first downward deflection of the QRS complex. The entire QRS complex marks the onset of ventricular depolarisation, which immediately precedes ventricular contraction (systole). Therefore, it signifies the **beginning of systole**. This is the best fit among the options and matches with **I. Beginning of systole**.

- **C. QRS complex:** This complex represents the depolarisation of the ventricles, which initiates ventricular contraction. This matches with **IV. Depolarisation of ventricles**.

- **D. T-wave:** This represents the return of the ventricles from the excited to the normal state (repolarisation). The end of the T-wave marks the end of systole. This matches with **II. Repolarisation of ventricles**.

Step 3: Final Answer
Based on the matching:
A matches with III.
B matches with I.
C matches with IV.
D matches with II.
This combination corresponds to **A-III, B-I, C-IV, D-II**, which is option (3).

> **Quick Tip**
>
> Remember the sequence: **P**olarization of **A**tria (P-wave), then **QRS** for ventricular de**P**olarization, and finally **T** for ventricular re**P**olarization. Depolarization leads to contraction, Repolarization leads to relaxation.

---

161. Match List I with List II.

List I		**List II**	
A.	**CCK**	**I.**	**Kidney**
B.	**GIP**	**II.**	**Heart**
C.	**ANF**	**III.**	**Gastric gland**
D.	**ADH**	**IV.**	**Pancreas**

Choose the correct answer from the options given below:

(A) A-II, B-IV, C-I, D-III
(B) A-IV, B-II, C-III, D-I
(C) A-IV, B-III, C-II, D-I
(D) A-III, B-II, C-IV, D-I

Correct Answer: (3) A-IV, B-III, C-II, D-I

Solution:

Step 1: Understanding the Question:
This question requires matching the hormones in List I with their primary site of action or secretion organ in List II.

Step 2: Detailed Explanation:
Let's analyze each hormone and its corresponding function/organ:

- **A. CCK (Cholecystokinin):** This is a gastrointestinal hormone secreted by the duodenum. It acts on the pancreas to stimulate the secretion of pancreatic enzymes and on the gallbladder to stimulate the release of bile. In the given options, Pancreas is listed. So, A matches with IV.

- **B. GIP (Gastric Inhibitory Peptide):** Also a gastrointestinal hormone, GIP inhibits gastric secretion and motility. Thus, its target is the gastric glands in the stomach. So, B matches with III.

- **C. ANF (Atrial Natriuretic Factor):** This hormone is secreted by the atrial walls of the heart in response to high blood pressure. It causes vasodilation and excretion of $Na+$ by the kidneys, thus lowering blood pressure. The source organ is the Heart. So, C matches with II.

- **D. ADH (Antidiuretic Hormone or Vasopressin):** This hormone is released from the posterior pituitary but synthesized in the hypothalamus. It acts on the distal convoluted tubule and collecting ducts of the nephrons in the kidney to increase water reabsorption. Its target organ is the Kidney. So, D matches with I.

Combining the correct matches: A-IV, B-III, C-II, D-I.

Step 3: Final Answer:
Based on the analysis, the correct combination is A-IV, B-III, C-II, D-I, which is option (3).

> **Quick Tip**
>
> For hormone questions, focus on both the gland of secretion and the target organ. CCK and GIP are gut hormones. ANF is from the heart (Atrial). ADH acts on the kidneys to conserve water (Anti-diuresis).

---

162. Match List I with List II.

List I		**List II**	
A.	**Gene 'a'**	**I.**	**β-galactosidase**
B.	**Gene 'y'**	**II.**	**Transacetylase**
C.	**Gene 'i'**	**III.**	**Permease**
D.	**Gene 'z'**	**IV.**	**Repressor protein**

Choose the correct answer from the options given below:

(A) A-III, B-IV, C-I, D-II
(B) A-III, B-I, C-IV, D-II
(C) A-II, B-I, C-IV, D-III
(D) A-II, B-III, C-IV, D-I

Correct Answer: (4) A-II, B-III, C-IV, D-I

Solution:

Step 1: Understanding the Question:
This question asks to match the genes of the lac operon (List I) with the proteins they code for (List II).

Step 2: Detailed Explanation:
The lac operon in E. coli consists of several genes involved in lactose metabolism. Let's identify the function of each gene listed:

- **Gene 'i' (Regulator gene):** This gene codes for the repressor protein. The repressor protein binds to the operator region to switch off the operon in the absence of lactose. So, C matches with IV.

- **Gene 'z' (Structural gene):** This gene codes for the enzyme β-galactosidase, which hydrolyzes lactose into glucose and galactose. So, D matches with I.

- **Gene 'y' (Structural gene):** This gene codes for the enzyme permease, which increases the permeability of the cell to lactose, allowing it to enter the cell. So, B matches with III.

- **Gene 'a' (Structural gene):** This gene codes for the enzyme transacetylase. So, A matches with II.

Combining the correct matches:
A $\rightarrow$ II (Transacetylase)
B $\rightarrow$ III (Permease)
C $\rightarrow$ IV (Repressor protein)
D $\rightarrow$ I (β-galactosidase)
This corresponds to the combination A-II, B-III, C-IV, D-I.

Step 3: Final Answer:
The correct matching is given in option (4).

Quick Tip

Remember the order and function of genes in the lac operon: i-p-o-z-y-a. 'i' is the regulator gene (repressor). 'z', 'y', and 'a' are the structural genes coding for β-galactosidase, Permease, and Transacetylase, respectively.

---

163. Vital capacity of lung is ________.

(A) IRV + ERV + TV – RV
(B) IRV + ERV + TV
(C) IRV + ERV
(D) IRV + ERV + TV + RV

Correct Answer: (2) IRV + ERV + TV

Solution:

Step 1: Understanding the Question
The question asks for the correct formula for the Vital Capacity (VC) of the lungs.

Step 2: Key Formula or Approach
We need to define the relevant lung volumes:

- **Tidal Volume (TV):** Volume of air inspired or expired during a normal respiration (~500 ml).
- **Inspiratory Reserve Volume (IRV):** Additional volume of air a person can inspire by a forcible inspiration (~2500-3000 ml).
- **Expiratory Reserve Volume (ERV):** Additional volume of air a person can expire by a forcible expiration (~1000-1100 ml).
- **Residual Volume (RV):** Volume of air remaining in the lungs even after a forcible expiration (~1100-1200 ml).

Vital Capacity (VC) is defined as the maximum volume of air a person can breathe out after a forced inspiration. It represents the maximum amount of exchangeable air.

Step 3: Detailed Explanation
The calculation for Vital Capacity is the sum of the volumes that can be moved in and out of the lungs:

$$\text{VC} = \text{ERV} + \text{TV} + \text{IRV}$$

Let's analyze the given options:
(1) IRV + ERV + TV – RV: This is incorrect.
(2) IRV + ERV + TV: This matches the correct definition of Vital Capacity.

(3) IRV + ERV: This is incorrect; it omits the Tidal Volume. This sum is sometimes referred to just as the sum of reserves, not the full vital capacity.
(4) IRV + ERV + TV + RV: This sum is equal to the Total Lung Capacity (TLC), not the Vital Capacity.

Step 4: Final Answer
The correct formula for Vital Capacity is the sum of Inspiratory Reserve Volume, Expiratory Reserve Volume, and Tidal Volume. Therefore, option (2) is the correct answer.

> **Quick Tip**
>
> Remember that "Vital Capacity" is the total 'usable' or 'exchangeable' volume of air. "Residual Volume" is what you can't exchange. Total Lung Capacity = Vital Capacity + Residual Volume.

---

164. Select the correct group/set of Australian Marsupials exhibiting adaptive radiation.

(A) Mole, Flying squirrel, Tasmanian tiger cat
(B) Lemur, Anteater, Wolf
(C) Tasmanian wolf, Bobcat, Marsupial mole
(D) Numbat, Spotted cuscus, Flying phalanger

Correct Answer: (4) Numbat, Spotted cuscus, Flying phalanger

Solution:

Step 1: Understanding the Question:
The question asks to identify the option that contains only Australian marsupials, which are a classic example of adaptive radiation. Adaptive radiation is the diversification of a group of organisms into forms filling different ecological niches.

Step 2: Detailed Explanation:
Let's analyze the organisms listed in each option:

- **(1) Mole, Flying squirrel, Tasmanian tiger cat:** The Mole and Flying squirrel are names of placental mammals. While there are marsupial equivalents (Marsupial mole, Flying phalanger) that show convergent evolution, this list mixes placental mammals with a marsupial (Tasmanian tiger cat was a marsupial). Thus, this is not a correct set.

- **(2) Lemur, Anteater, Wolf:** Lemur is a primate, and the Anteater and Wolf are other placental mammals. This list contains no marsupials.

- **(3) Tasmanian wolf, Bobcat, Marsupial mole:** Tasmanian wolf (Thylacine) and Marsupial mole are Australian marsupials. However, the Bobcat is a placental mammal (a species of wild cat found in North America). This list is a mix.

- **(4) Numbat, Spotted cuscus, Flying phalanger:** The Numbat (also known as the marsupial anteater), the Spotted cuscus, and the Flying phalanger are all Australian marsupials. They evolved from a common marsupial ancestor and adapted to different lifestyles and diets within Australia. This is a perfect example of a set of animals from a single adaptive radiation event.

Step 3: Final Answer:
Option (4) is the only one that lists a group consisting entirely of Australian marsupials.

> Quick Tip
>
> Be familiar with the classic examples of adaptive radiation: Darwin's finches in the Galapagos and Australian marsupials. For marsupials, learn to distinguish them from their placental counterparts that show convergent evolution (e.g., Marsupial mole vs. Placental mole, Tasmanian wolf vs. Placental wolf).

---

165. Match List I with List II.

	List I		**List II**
A.	Ringworm	I.	Haemophilus influenzae
B.	Filariasis	II.	Trichophyton
C.	Malaria	III.	Wuchereria bancrofti
D.	Pneumonia	IV.	Plasmodium vivax

Choose the correct answer from the options given below:

(A) A-III, B-II, C-I, D-IV
(B) A-III, B-II, C-IV, D-I
(C) A-II, B-III, C-IV, D-I
(D) A-II, B-III, C-I, D-IV

Correct Answer: (3) A-II, B-III, C-IV, D-I

Solution:

Step 1: Understanding the Question
This question requires matching common diseases (List I) with their causative pathogens (List II).

Step 2: Detailed Explanation
Let's match each disease with its pathogen:

- **A. Ringworm:** Despite its name, ringworm is not caused by a worm. It is a common fungal infection of the skin. Genera of fungi like **Trichophyton**, Microsporum, and Epidermophyton cause ringworm. This matches with **II**.

- **B. Filariasis (Elephantiasis):** This is a parasitic disease caused by filarial worms (helminths). The most common causative agent is **Wuchereria bancrofti**. This matches with **III**.

- **C. Malaria:** This is a mosquito-borne infectious disease caused by a protozoan parasite of the genus *Plasmodium*. **Plasmodium vivax** is one of the species that causes malaria in humans. This matches with **IV**.

- **D. Pneumonia:** This is an infection that inflames the air sacs in one or both lungs. It can be caused by bacteria, viruses, or fungi. A common bacterial cause is **Haemophilus influenzae** (another is *Streptococcus pneumoniae*). This matches with **I**.

Step 3: Final Answer
Based on the matching:
A matches with II.
B matches with III.
C matches with IV.
D matches with I.
This combination corresponds to **A-II, B-III, C-IV, D-I**, which is option (3).

> Quick Tip
>
> For disease questions, create a table with four columns: Disease Name, Causative Agent, Type of Pathogen (Virus, Bacteria, Protozoa, Fungus, Helminth), and Mode of Transmission/Vector. This organized information is crucial for exams.

---

166. Match List I with List II.

List I (Type of Joint)		**List II (Found between)**	
A.	**Cartilaginous Joint**	**I.**	**Between flat skull bones**
B.	**Ball and Socket Joint**	**II.**	**Between adjacent vertebrae in vertebral column**
C.	**Fibrous Joint**	**III.**	**Between carpal and metacarpal of thumb**
D.	**Saddle Joint**	**IV.**	**Between Humerus and Pectoral girdle**

Choose the correct answer from the options given below:

(A) A-I, B-IV, C-III, D-II
(B) A-II, B-IV, C-III, D-I
(C) A-III, B-I, C-II, D-IV
(D) A-II, B-IV, C-I, D-III

Correct Answer: (4) A-II, B-IV, C-I, D-III

Solution:

Step 1: Understanding the Question:
The question requires matching the type of joint given in List I with its correct location in the human body from List II.

Step 2: Detailed Explanation:
Let's analyze each joint type and its location:

- **A. Cartilaginous Joint:** These joints have bones connected by cartilage and allow limited movement. A classic example is the joints between adjacent vertebral bodies in the spinal column, which are connected by intervertebral discs (fibrocartilage). So, A matches with II.

- **B. Ball and Socket Joint:** This is a type of synovial joint that allows for a wide range of motion (multiaxial). The shoulder joint (between the head of the Humerus and the glenoid cavity of the Pectoral girdle) and the hip joint are prime examples. So, B matches with IV.

- **C. Fibrous Joint:** These joints are connected by dense fibrous connective tissue and are typically immovable (synarthrosis). The sutures between the flat bones of the skull are a key example. So, C matches with I.

- **D. Saddle Joint:** This is another type of synovial joint that allows movement in two planes (biaxial). The carpometacarpal joint between the carpal bone (trapezium) and the first metacarpal of the thumb is the classic example in the human body. So, D matches with III.

Combining the correct matches: A-II, B-IV, C-I, D-III.

Step 3: Final Answer:
Based on the matching, the correct combination is A-II, B-IV, C-I, D-III, which is option (4).

Quick Tip

To master joints, create a table with three columns: Joint Type (Fibrous, Cartilaginous, Synovial), Degree of Movement (Immovable, Slightly Movable, Freely Movable), and a key Example (Skull Sutures, Vertebrae, Shoulder/Knee). For synovial joints, learn the subtypes (ball-and-socket, hinge, pivot, saddle, etc.) and their specific examples.

---

167. Given below are statements: one is labelled as Assertion A and the other is labelled as Reason R.

Assertion A: Nephrons are of two types: Cortical & Juxta medullary, based on their relative position in cortex and medulla.

Reason R: Juxta medullary nephrons have short loop of Henle whereas, cortical nephrons have longer loop of Henle.

In the light of the above statements, choose the correct answer from the options given below:

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (1) A is true but R is false.

Solution:

Step 1: Understanding the Question

The question asks us to evaluate an Assertion and a Reason related to the types of nephrons in the human kidney.

Step 2: Detailed Explanation

Analysis of Assertion A:

The assertion states that nephrons are classified into two types, cortical and juxtamedullary, based on their location. This is correct. Cortical nephrons are more numerous ($\sim$85%) and have their renal corpuscles in the superficial part of the cortex. Juxtamedullary nephrons have their renal corpuscles deep in the cortex, near the medulla. Thus, **Assertion A is true**.

Analysis of Reason R:

The reason describes the length of the loop of Henle in these two types of nephrons. However, it states that juxtamedullary nephrons have a short loop of Henle and cortical nephrons have a long loop. This is incorrect. **Juxtamedullary nephrons have a very long loop of Henle** that extends deep into the medulla. This long loop is crucial for creating the concentration gradient necessary for producing concentrated urine. **Cortical nephrons have a short loop of Henle** that barely dips into the medulla. The statement reverses these facts. Thus, **Reason R is false**.

Step 3: Final Answer

Since the Assertion is true and the Reason is false, the correct option is (1).

Quick Tip

Remember: "Juxtamedullary" means "next to the medulla". These nephrons are specialized for water conservation, which requires a long loop of Henle to create a strong osmotic gradient deep in the medulla.

---

168. Which one of the following techniques does not serve the purpose of early diagnosis of a disease for its early treatment?

(A) Polymerase Chain Reaction (PCR) technique
(B) Enzyme Linked Immuno-Sorbent Assay (ELISA) technique
(C) Recombinant DNA Technology
(D) Serum and Urine analysis

Correct Answer: (4) Serum and Urine analysis

Solution:

Step 1: Understanding the Question:
The question asks to identify which of the given techniques is generally not used for *early* diagnosis of a disease. Early diagnosis means detecting the disease at a very initial stage, often before symptoms appear.

Step 2: Detailed Explanation:
Let's evaluate each technique's application in early diagnosis:

- **Polymerase Chain Reaction (PCR):** PCR can amplify a specific DNA or RNA sequence. This allows the detection of very small quantities of a pathogen's genetic material (like a virus) long before the body mounts a significant immune response or symptoms develop. It is a cornerstone of early molecular diagnosis.

- **Enzyme Linked Immuno-Sorbent Assay (ELISA):** ELISA is based on the antigen-antibody interaction. It can detect either the presence of antigens from a pathogen or the antibodies produced by the body in response to an infection. It is very sensitive and widely used for early diagnosis of diseases like HIV.

- **Recombinant DNA Technology:** This technology allows the creation of DNA probes and other tools that can detect specific gene mutations (e.g., in cancer or genetic disorders) or identify pathogenic DNA/RNA. It is a powerful tool for early diagnosis.

- **Serum and Urine analysis:** This is a conventional diagnostic method. It typically measures levels of metabolites, ions, proteins, or cells in blood serum or urine. While very useful, these parameters often change significantly only after the disease has progressed to a certain stage and caused physiological changes. Therefore, it is generally not considered a technique for *very early* diagnosis compared to molecular methods that detect the causative agent or genetic defect directly.

Conclusion: PCR, ELISA, and Recombinant DNA Technology are modern molecular techniques that enable early detection. Serum and urine analysis is a more traditional method that often detects a disease at a later stage.

Step 3: Final Answer:
Serum and Urine analysis does not serve the purpose of early diagnosis as effectively as the other listed techniques. Hence, option (4) is the correct answer.

> **Quick Tip**
>
> For questions about diagnostic techniques, remember that molecular methods (PCR, rDNA tech) and sensitive immunological assays (ELISA) are key for "early diagnosis" because they can detect minute amounts of pathogens or genetic markers, often before symptoms are visible.

---

169. Which one of the following common sexually transmitted diseases is completely curable when detected early and treated properly?

(A) Hepatitis-B
(B) HIV Infection
(C) Genital herpes
(D) Gonorrhoea

Correct Answer: (4) Gonorrhoea

Solution:

Step 1: Understanding the Question
The question asks to identify which of the listed sexually transmitted diseases (STDs) is completely curable. The key distinction is between curable (usually bacterial) and manageable/non-curable (usually viral) STDs.

Step 2: Detailed Explanation
Let's analyze the given options:

- **(1) Hepatitis-B:** This is a viral infection that affects the liver, caused by the Hepatitis B virus (HBV). While it can be managed with antiviral medications, there is no complete cure for chronic Hepatitis-B.

- **(2) HIV Infection:** This is caused by the Human Immunodeficiency Virus (HIV). With Antiretroviral therapy (ART), the virus can be suppressed to undetectable levels, allowing for a long and healthy life, but it cannot be completely eliminated from the body. It is not curable.

- **(3) Genital herpes:** This is caused by the Herpes Simplex Virus (HSV). Antiviral medications can manage outbreaks and reduce transmission, but they cannot eradicate the virus from the body. It is a lifelong infection and not curable.

- **(4) Gonorrhoea:** This is a bacterial infection caused by *Neisseria gonorrhoeae*. As a bacterial infection, it is completely curable with a course of appropriate antibiotics, especially when diagnosed and treated early.

Step 3: Final Answer
Among the given options, only Gonorrhoea is a bacterial STD and is therefore completely curable with antibiotics. The others are viral infections that are currently not curable. Hence, option (4) is the correct answer.

> **Quick Tip**
>
> A general rule for STDs in exams: bacterial infections (like Syphilis, Gonorrhoea, Chlamydia) are generally curable with antibiotics. Viral infections (like HIV, Herpes, Hepatitis-B, HPV) are generally not curable, though they can be managed.

---

170. Given below are two statements:
Statement I: Electrostatic precipitator is most widely used in thermal power plant.
Statement II: Electrostatic precipitator in thermal power plant removes ionising radiations.
In the light of the above statements, choose the most appropriate answer from the options given below:

(A) Statement I is correct but Statement II is incorrect.
(B) Statement I incorrect but Statement II is correct.
(C) Both Statement I and Statement II are correct.
(D) Both Statement I and Statement II are incorrect.

Correct Answer: (1) Statement I is correct but Statement II is incorrect.

Solution:

Step 1: Understanding the Question:
The question asks to evaluate two statements regarding the function and use of electrostatic

precipitators in thermal power plants.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** Thermal power plants, which burn fossil fuels like coal, are a major source of particulate matter (ash) pollution. The Electrostatic Precipitator (ESP) is a device that removes suspended dust particles from a gas or exhaust stream by applying a high-voltage electrostatic charge. They are highly efficient (can remove over 99% of particulate matter) and are indeed the most widely used method for controlling particulate emissions from thermal power plants. Thus, Statement I is correct.

- **Analysis of Statement II:** This statement claims that an ESP removes ionising radiations. This is incorrect. An ESP is designed to remove physical **particulate matter** (like dust, smoke, and ash). It works by charging these particles and collecting them on plates. Ionising radiations (like X-rays, gamma rays) are forms of electromagnetic energy, not physical particles in this context. They are not affected by the electrostatic fields in an ESP. Removing radiation requires shielding with dense materials like lead or concrete. Thus, Statement II is incorrect.

Step 3: Final Answer:
Statement I is correct, but Statement II is incorrect. Therefore, option (1) is the correct answer.

> Quick Tip
>
> Associate pollution control devices with the type of pollutant they remove. Electrostatic precipitators and scrubbers remove particulate matter. Catalytic converters in cars reduce gaseous pollutants like CO, NOx, and unburnt hydrocarbons. ESPs do NOT deal with radiation.

---

171. Which of the following is not a cloning vector?

(A) pBR322
(B) Probe
(C) BAC
(D) YAC

Correct Answer: (2) Probe

Solution:

Step 1: Understanding the Question
The question asks to identify which of the given options is not a cloning vector.

Step 2: Detailed Explanation
A **cloning vector** is a small piece of DNA that can be stably maintained in an organism, and into which a foreign DNA fragment can be inserted for cloning purposes. It must have features like an origin of replication, selectable markers, and cloning sites. Let's analyze the options:

- **(1) pBR322:** This is one of the first widely used E. coli cloning vectors. It is a plasmid and a classic example of a cloning vector.
- **(2) Probe:** A DNA probe is a single-stranded DNA or RNA fragment that is complementary to a specific DNA sequence of interest. It is labeled (e.g., with a radioactive or fluorescent tag) and used to detect the presence of that sequence in a sample. It is a detection tool, not a vehicle for carrying and replicating DNA.
- **(3) BAC (Bacterial Artificial Chromosome):** This is a cloning vector based on the F-plasmid of E. coli. It is used to clone large DNA fragments (100-300 kb) in bacteria.
- **(4) YAC (Yeast Artificial Chromosome):** This is a cloning vector that can carry very large DNA fragments (up to a million base pairs) and replicate them in yeast cells.

Step 3: Final Answer
pBR322, BAC, and YAC are all types of cloning vectors used in genetic engineering. A probe is a tool for detection, not cloning. Therefore, "Probe" is not a cloning vector. Option (2) is the correct answer.

> **Quick Tip**
> Remember the function: a **vector** is a **vehicle** used to carry genetic material into a cell. A **probe** is a **detector** used to find a specific gene. Don't confuse the vehicle with the detector.

---

172. Radial symmetry is NOT found in adults of phylum

(A) Coelenterata
(B) Echinodermata
(C) Ctenophora
(D) Hemichordata

Correct Answer: (4) Hemichordata

Solution:

Step 1: Understanding the Question:
The question asks to identify the animal phylum whose adult members do not exhibit radial symmetry.

Step 2: Detailed Explanation:
Let's examine the symmetry of the adult forms in each phylum listed:

- **(1) Coelenterata (Cnidaria):** Animals like jellyfish, sea anemones, and corals belong to this phylum. Their adults typically have a body plan with parts arranged around a central axis, which is characteristic of **radial symmetry**.

- **(2) Echinodermata:** This phylum includes starfish, sea urchins, and sea cucumbers. The adult echinoderms are known for their characteristic **pentamerous radial symmetry** (body parts arranged in fives around a central axis). Interestingly, their larvae are bilaterally symmetric.

- **(3) Ctenophora:** Commonly known as comb jellies, these animals exhibit **biradial symmetry**, which is a type of radial symmetry where the body can be divided into two equal halves by only two planes.

- **(4) Hemichordata:** This phylum consists of worm-like marine animals such as *Balanoglossus* (acorn worm). These animals have a distinct anterior and posterior end, as well as dorsal and ventral sides. Their body can be divided into two equal left and right halves by only one plane. This is characteristic of **bilateral symmetry**.

Step 3: Final Answer:
Adults of the phylum Hemichordata exhibit bilateral symmetry, not radial symmetry. Therefore, option (4) is the correct answer.

Quick Tip

Symmetry is a fundamental concept in classifying animal phyla. Remember the key examples: Cnidaria and adult Echinodermata are radially symmetric. Most other advanced phyla, including Annelida, Arthropoda, Mollusca, Hemichordata, and Chordata, are bilaterally symmetric.

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173. Match List I with List II.

List I (Cells)		**List II (Secretion)**	
A.	Peptic cells	I.	Mucus
B.	Goblet cells	II.	Bile juice
C.	Oxyntic cells	III.	Proenzyme pepsinogen
D.	Hepatic cells	IV.	HCl and intrinsic factor for absorption of vitamin B_{12}

Choose the correct answer from the options given below:

(A) A-III, B-I, C-IV, D-II
(B) A-II, B-IV, C-I, D-III
(C) A-IV, B-III, C-II, D-I
(D) A-II, B-I, C-III, D-IV

Correct Answer: (1) A-III, B-I, C-IV, D-II

Solution:

Step 1: Understanding the Question
The question requires matching different types of cells from the digestive system (List I) with their respective secretions (List II).

Step 2: Detailed Explanation
Let's match each cell type to its secretion:

- **A. Peptic cells:** Also known as chief cells or zymogenic cells, these are found in the gastric glands of the stomach. They secrete the inactive proenzyme **pepsinogen**. This matches with **III**.

- **B. Goblet cells:** These are found throughout the epithelial lining of the gastrointestinal tract. They secrete **mucus**, which lubricates and protects the lining. This matches with **I**.

- **C. Oxyntic cells:** Also known as parietal cells, these are also found in the gastric glands. They secrete **hydrochloric acid (HCl)** and **intrinsic factor**, which is essential for the absorption of vitamin B_{12}. This matches with **IV**.

- **D. Hepatic cells:** These are the main cells of the liver (hepatocytes). They produce and secrete **bile juice**, which is important for the emulsification of fats. This matches with **II**.

Step 3: Final Answer
Based on the matching:
A matches with III.
B matches with I.
C matches with IV.
D matches with II.
This combination corresponds to **A-III, B-I, C-IV, D-II**, which is option (1).

Quick Tip

For the stomach glands, remember: **P**eptic cells secrete **P**epsinogen. **P**arietal (Oxyntic) cells secrete HCl and intrinsic factor. Goblet cells are the "goblets" full of mucus.

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174. Once the undigested and unabsorbed substances enter the caecum, their backflow is prevented by-

(A) Gastro - oesophageal sphincter
(B) Pyloric sphincter
(C) Sphincter of Oddi
(D) Ileo - caecal valve

Correct Answer: (4) Ileo - caecal valve

Solution:

Step 1: Understanding the Question:
The question asks to identify the structure that prevents the backward movement of contents from the caecum (the beginning of the large intestine) into the small intestine.

Step 2: Detailed Explanation:
Let's review the function of each sphincter/valve listed:

- **Gastro-oesophageal sphincter:** Located between the esophagus and the stomach. It prevents the backflow of stomach contents (acidic chyme) into the esophagus.
- **Pyloric sphincter:** Located between the stomach and the duodenum (the first part of the small intestine). It regulates the passage of chyme from the stomach into the small intestine.
- **Sphincter of Oddi:** Guards the opening of the common hepato-pancreatic duct into the duodenum. It controls the flow of bile and pancreatic juice into the small intestine.
- **Ileo-caecal valve:** Located at the junction of the ileum (the last part of the small intestine) and the caecum (the first part of the large intestine). Its primary function is to prevent the reflux of colonic contents into the ileum.

Therefore, the ileo-caecal valve is the structure that prevents the backflow of substances from the caecum.

Step 3: Final Answer:
The correct structure is the Ileo-caecal valve, which is option (4).

> Quick Tip
>
> Remember the path of food through the digestive system and the names of the sphincters that regulate flow at each major junction: Esophagus $\rightarrow$ Gastro-oesophageal sphincter $\rightarrow$ Stomach $\rightarrow$ Pyloric sphincter $\rightarrow$ Small Intestine (Duodenum, Jejunum, Ileum) $\rightarrow$ Ileo-caecal valve $\rightarrow$ Large Intestine (Caecum, Colon, Rectum).

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175. Given below are two statements:
Statement I: In prokaryotes, the positively charged DNA is held with some negatively charged proteins in a region called nucleoid.
Statement II: In eukaryotes, the negatively charged DNA is wrapped around the positively charged histone octamer to form nucleosome.
In the light of the above statements, choose the correct answer from the options given below:

(A) Statement I is correct but Statement II is false.
(B) Statement I incorrect but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (2) Statement I incorrect but Statement II is true.

Solution:

Step 1: Understanding the Question:
The question presents two statements about DNA packaging in prokaryotes and eukaryotes and asks for an evaluation of their correctness.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** "In prokaryotes, the positively charged DNA is held with some negatively charged proteins in a region called nucleoid." This statement contains factual errors regarding the charges. DNA, due to its phosphate backbone (PO_4^{3-}), is **negatively charged**. To package this negatively charged DNA, it is associated with **positively charged** proteins (non-histone proteins in prokaryotes). The statement incorrectly reverses these charges. Therefore, Statement I is incorrect.

- **Analysis of Statement II:** "In eukaryotes, the negatively charged DNA is wrapped around the positively charged histone octamer to form nucleosome." This statement is correct. Eukaryotic DNA is also **negatively charged**. It is wrapped around a core of eight histone proteins (a histone octamer). Histone proteins are rich in basic (positively charged) amino acids like lysine and arginine, which gives them a net positive charge. This positive charge allows them to bind tightly to the negatively charged DNA. This complex of DNA wrapped around a histone octamer is called a nucleosome, the basic unit of chromatin. Therefore, Statement II is true.

Step 3: Final Answer:
Statement I is incorrect, while Statement II is true. This corresponds to option (2).

> Quick Tip
>
> A fundamental concept in molecular biology is that DNA is always negatively charged due to its phosphate groups. For packaging, it must associate with positively charged proteins (histones in eukaryotes, non-histone proteins in prokaryotes). Getting these charges right is crucial.

---

176. Given below are two statements:
Statement I: Ligaments are dense irregular tissue.
Statement II: Cartilage is dense regular tissue.
In the light of the above statements, choose the correct answer from the options given below:

(A) Statement I is true but Statement II is false.
(B) Statement I is false but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (4) Both Statement I and Statement II are false.

Solution:

Step 1: Understanding the Question:
The question asks us to evaluate the correctness of two statements regarding types of connective tissues.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** Ligaments are fibrous connective tissues that connect bone to bone. They are composed of collagen fibres that are arranged in a parallel fashion to provide high tensile strength in one direction. This type of tissue is known as **dense regular connective tissue**, not dense irregular tissue. Dense irregular tissue has collagen fibres arranged in a non-parallel, random manner and is found in the dermis of the skin.

Therefore, Statement I is false.

- **Analysis of Statement II:** Cartilage is a type of specialized connective tissue. It is not classified as dense regular tissue. Dense regular tissue is a sub-category of connective tissue proper. Cartilage is distinct, characterized by cells called chondrocytes embedded in a firm, pliable matrix called chondrin. Therefore, Statement II is false.

Step 3: Final Answer:
Since both statements are incorrect, the correct option is (4).

Quick Tip

Remember the classification of connective tissues. Distinguish between Connective Tissue Proper (loose and dense), Specialized Connective Tissue (cartilage, bone, blood), etc. For dense connective tissue, remember: Tendons (muscle to bone) and Ligaments (bone to bone) are Dense Regular. The dermis is Dense Irregular.

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177. Given below are two statements:
Statement I: Low temperature preserves the enzyme in a temporarily inactive state whereas high temperature destroys enzymatic activity because proteins are denatured by heat.
Statement II: When the inhibitor closely resembles the substrate in its molecular structure and inhibits the activity of the enzyme, it is known as competitive inhibitor.
In the light of the above statements, choose the correct answer from the options given below:

(A) Statement I is true but Statement II is false.
(B) Statement I is false but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (3) Both Statement I and Statement II are true.

Solution:

Step 1: Understanding the Question:
The question presents two statements related to enzyme kinetics and asks to evaluate their correctness.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** This statement describes the effect of temperature on enzyme activity.
  - At low temperatures, enzymes become temporarily inactive because molecules have less kinetic energy, reducing the frequency of effective collisions between the enzyme and substrate. This inactivation is reversible. Preserving food by refrigeration is based on this principle.
  - At high temperatures (beyond the optimum), the thermal energy breaks the weak hydrogen bonds that maintain the enzyme's specific three-dimensional (tertiary) structure. This irreversible change in shape is called denaturation, and it leads to a permanent loss of enzymatic activity.
  Thus, Statement I is correct.

- **Analysis of Statement II:** This statement defines a competitive inhibitor.
  - A competitive inhibitor is a molecule that has a molecular structure very similar to the enzyme's actual substrate.
  - Because of this structural similarity, it can bind to the active site of the enzyme, competing with the substrate.
  - When the inhibitor is bound to the active site, the substrate cannot bind, and the enzyme's activity is inhibited.
  This is the precise definition of competitive inhibition. Thus, Statement II is correct.

Step 3: Final Answer:
Since both Statement I and Statement II are correct descriptions of principles in enzymology, option (3) is the correct choice.

Quick Tip

Remember the key differences in enzyme inhibition: Competitive inhibitors bind to the active site and resemble the substrate. Non-competitive inhibitors bind to an allosteric site and do not resemble the substrate. Both high temperature (denaturation) and extreme pH can irreversibly destroy enzyme function.

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178. Which of the following statements is correct?

(A) Presence of large amount of nutrients in water restricts 'Algal Bloom'.
(B) Algal Bloom decreases fish mortality.
(C) Eutrophication refers to increase in domestic sewage and waste water in lakes.
(D) Biomagnification refers to increase in concentration of the toxicant at successive trophic levels.

Correct Answer: (4) Biomagnification refers to increase in concentration of the toxicant at successive trophic levels.

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statement among the four options related to environmental issues.

Step 2: Detailed Explanation:
Let's analyze each statement:

- **(1) Presence of large amount of nutrients in water restricts 'Algal Bloom'.** This is incorrect. A large amount of nutrients (like nitrates and phosphates) in water bodies leads to excessive growth of planktonic algae, which is known as an algal bloom. The nutrients promote, not restrict, the bloom.

- **(2) Algal Bloom decreases fish mortality.** This is incorrect. Algal blooms cause a deterioration in water quality. When the algae die, they are decomposed by bacteria, a process that consumes large amounts of dissolved oxygen in the water. This depletion of oxygen (hypoxia or anoxia) leads to the death of fish and other aquatic animals, thus increasing fish mortality.

- **(3) Eutrophication refers to increase in domestic sewage and waste water in lakes.** This is an imprecise definition. Eutrophication is the process of nutrient enrichment of a water body that leads to excessive plant and algal growth. While domestic sewage is a major cause of cultural eutrophication, the term itself refers to the nutrient enrichment and its consequences, not just the addition of sewage.

- **(4) Biomagnification refers to increase in concentration of the toxicant at successive trophic levels.** This is the correct definition of biomagnification (also known as bioamplification or biological magnification). It occurs because certain toxic substances (like DDT, mercury) are absorbed by organisms and are not easily metabolized or excreted. As these organisms are consumed by others higher up the food chain, the toxicant becomes increasingly concentrated at each trophic level.

Step 3: Final Answer:
Statement (4) provides the correct definition of biomagnification. Therefore, it is the correct answer.

> **Quick Tip**
>
> Distinguish between key ecological terms: Eutrophication = nutrient enrichment. Algal Bloom = result of eutrophication. Biomagnification = concentration of toxins up the food chain. Bioaccumulation = concentration of a toxin in a single organism over its lifetime.

179. Given below are two statements:
Statement I: A protein is imagined as a line, the left end represented by first amino acid (C-terminal) and the right end represented by last amino acid (N-terminal)
Statement II: Adult human haemoglobin, consists of 4 subunits (two subunits of α type and two subunits of β type.)
In the light of the above statements, choose the correct answer from the options given below:

(A) Statement I is true but Statement II is false.
(B) Statement I is false but Statement II is true.
(C) Both Statement I and Statement II are true.
(D) Both Statement I and Statement II are false.

Correct Answer: (2) Statement I is false but Statement II is true.

Solution:

Step 1: Understanding the Question
The question asks to evaluate the correctness of two statements regarding the structure of proteins and hemoglobin.

Step 2: Detailed Explanation
Analysis of Statement I:
A protein is a polypeptide chain of amino acids. By convention, the sequence of amino acids in a protein is written starting from the amino-terminal end (N-terminus), which has a free amino group (-NH_2). The last amino acid in the chain is the carboxyl-terminal end (C-terminus), which has a free carboxyl group (-COOH). Statement I incorrectly reverses this convention, stating the first amino acid is the C-terminal and the last is the N-terminal. Therefore, **Statement I is false**.

Analysis of Statement II:
Adult human hemoglobin (HbA) is a globular protein responsible for oxygen transport. It has a quaternary structure, meaning it is composed of multiple polypeptide subunits. Specifically, it is a tetramer made of four subunits: two identical alpha (α) chains and two identical beta (β) chains. This structure is often denoted as $\alpha_2\beta_2$. Therefore, **Statement II is true**.

Step 3: Final Answer
Since Statement I is false and Statement II is true, the correct option is (2).

Quick Tip

Remember the convention for proteins: the chain starts at the N-terminus and ends at the C-terminus. Think of it alphabetically: N comes before C in the reverse alphabet, but in biology, N-terminus is the "start" and C-terminus is the "end".

180. In which blood corpuscles, the HIV undergoes replication and produces progeny viruses?

(A) Basophils
(B) Eosinophils
(C) T_H cells
(D) B-lymphocytes

Correct Answer: (3) T_H cells

Solution:

Step 1: Understanding the Question:
The question asks to identify the specific type of blood cell where the Human Immunodeficiency Virus (HIV) replicates.

Step 2: Detailed Explanation:
HIV is a retrovirus that primarily targets cells of the immune system. The virus has surface glycoproteins (gp120) that bind to a specific receptor called CD4, which is present on the surface of certain immune cells.

- **T_H cells (Helper T-lymphocytes):** These cells have a high density of CD4 receptors on their surface, making them the primary target for HIV infection. Once inside a T_H cell, HIV uses its enzyme, reverse transcriptase, to create DNA from its RNA genome. This viral DNA is then integrated into the host cell's DNA. The infected T_H cell is then forced to produce new virus particles, becoming a "virus factory".

- **B-lymphocytes:** These cells are responsible for producing antibodies. They are not the primary target for HIV replication.

- **Basophils and Eosinophils:** These are types of granulocytes involved in allergic reactions and parasitic infections, respectively. They are not the main target cells for HIV.

The progressive destruction of T_H cells by HIV leads to a weakened immune system, which is characteristic of Acquired Immuno Deficiency Syndrome (AIDS).

Step 3: Final Answer:
HIV replicates within Helper T-cells (T_H cells). Therefore, option (3) is the correct answer.

Quick Tip

Remember that HIV stands for Human Immunodeficiency Virus. It targets the "helpers" of the immune system, the Helper T-cells (T_H cells), which are also known as CD4+ T-cells because they have the CD4 receptor that the virus uses to enter.

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181. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R.
Assertion A: Endometrium is necessary for implantation of blastocyst.
Reason R: In the absence of fertilization, the corpus luteum degenerates that causes disintegration of endometrium.
In the light of the above statements, choose the correct answer from the options given below:

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (4) Both A and R are true but R is NOT the correct explanation of A.

Solution:

Step 1: Understanding the Question:
This question requires evaluating an Assertion and a Reason related to the human menstrual cycle and implantation. We need to determine if both statements are true and if the Reason correctly explains the Assertion.

Step 2: Detailed Explanation:

- **Analysis of Assertion (A):** "Endometrium is necessary for implantation of blastocyst." The endometrium is the inner lining of the uterus, which becomes thick, vascularized, and rich in glands during the secretory phase of the menstrual cycle, preparing it for pregnancy. The blastocyst (early embryo) embeds itself into this receptive endometrium to establish pregnancy. Without a properly developed endometrium, implantation cannot occur. Thus, Assertion A is true.

- **Analysis of Reason (R):** "In the absence of fertilization, the corpus luteum degenerates that causes disintegration of endometrium." After ovulation, the remnant of the ovarian follicle develops into the corpus luteum, which secretes progesterone. Progesterone maintains the endometrium. If fertilization does not occur, the corpus luteum degenerates after about 10-12 days. The resulting sharp drop in progesterone levels leads to the breakdown and shedding of the endometrium, which is known as menstruation. Thus, Reason R is also true.

- **Correlation between A and R:** Both statements are individually true. However, Reason R explains what happens in the *absence* of fertilization (menstruation). Assertion A describes a condition necessary for a successful pregnancy (implantation). The reason for the endometrium's necessity for implantation (Assertion A) is that it provides structural support, nourishment, and a blood supply to the developing embryo. Reason R, while related to the same cycle, describes the mechanism of menstruation, not the function of the endometrium during implantation. Therefore, R is not the correct explanation for A.

Step 3: Final Answer:
Both Assertion A and Reason R are true statements, but R does not correctly explain A. Therefore, option (4) is the correct answer.

Quick Tip

In Assertion-Reason questions, first check the validity of each statement independently. If both are true, then ask "Why?" for the Assertion. If the Reason answers that "Why?", it's the correct explanation. Here, "Why is endometrium necessary for implantation?" is not answered by "Because corpus luteum degenerates without fertilization".

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182. Match List I with List II.

List I		**List II**	
A.	**Taenia**	**I.**	**Nephridia**
B.	**Paramoecium**	**II.**	**Contractile vacuole**
C.	**Periplaneta**	**III.**	**Flame cells**
D.	**Pheretima**	**IV.**	**Urecose gland**

Choose the correct answer from the options give below:

(A) A-III, B-II, C-IV, D-I
(B) A-II, B-I, C-IV, D-III
(C) A-I, B-II, C-III, D-IV
(D) A-I, B-II, C-IV, D-III

Correct Answer: (1) A-III, B-II, C-IV, D-I

Solution:

Step 1: Understanding the Question:
The question requires matching the organisms in List I with their corresponding excretory or osmoregulatory structures in List II.

Step 2: Detailed Explanation:
Let's analyze each organism and its structure:

- **A. Taenia (Tapeworm):** It is a platyhelminth (flatworm). The excretory structures in flatworms are specialized cells called flame cells (or protonephridia). So, A matches with III.

- **B. Paramoecium:** It is a freshwater protozoan. For osmoregulation (excretion of excess water), it uses a specialized organelle called the contractile vacuole. So, B matches with II.

- **C. Periplaneta (Cockroach):** It is an insect. The primary excretory organs are Malpighian tubules. However, urecose glands, located in the mushroom gland of male cockroaches, also function in excretion by storing uric acid. In the given options, urecose gland is listed. So, C matches with IV.

- **D. Pheretima (Earthworm):** It is an annelid. The excretory organs in earthworms are coiled tubular structures called nephridia. So, D matches with I.

Combining the correct matches: A-III, B-II, C-IV, D-I.

Step 3: Final Answer:
Based on the analysis, the correct combination is A-III, B-II, C-IV, D-I, which corresponds to option (1).

Quick Tip

For matching questions in biology, focus on the phylum/class of the organism first. Often, specific excretory structures are characteristic of entire groups (e.g., flame cells in Platyhelminthes, nephridia in Annelida).

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183. Match List I with List II.

List I (Interacting species)		**List II (Name of Interaction)**	
A.	A Leopard and a Lion in a forest/grassland	I.	Competition
B.	A Cuckoo laying egg in a Crow's nest	II.	Brood parasitism
C.	Fungi and root of a higher plant in Mycorrhizae	III.	Mutualism
D.	A cattle egret and a Cattle in a field	IV.	Commensalism

Choose the correct answer from the options given below:

(A) A-III, B-IV, C-I, D-II
(B) A-II, B-III, C-I, D-IV
(C) A-I, B-II, C-III, D-IV

(D) A-I, B-II, C-IV, D-III

Correct Answer: (3) A-I, B-II, C-III, D-IV

Solution:

Step 1: Understanding the Question
The question requires matching specific examples of species interactions (List I) with the correct ecological term for that interaction (List II).

Step 2: Detailed Explanation
Let's analyze each example:

- **A. A Leopard and a Lion in a forest/grassland:** Both are large predators that may hunt for the same prey (e.g., deer, zebra). Since they utilize the same limited resource, they are in **Competition** with each other. This is an example of interspecific competition. This matches with **I**.

- **B. A Cuckoo laying egg in a Crow's nest:** The cuckoo lays its eggs in the nest of another bird species (the host, like a crow), which then unknowingly raises the cuckoo chick, often at the expense of its own offspring. This is a classic example of **Brood parasitism**. This matches with **II**.

- **C. Fungi and root of a higher plant in Mycorrhizae:** Mycorrhiza is a symbiotic association between a fungus and the roots of a vascular plant. The fungus helps the plant absorb nutrients (like phosphorus) from the soil, and the plant provides the fungus with carbohydrates. Both partners benefit. This is **Mutualism**. This matches with **III**.

- **D. A cattle egret and a Cattle in a field:** Cattle egrets are birds that follow grazing cattle. As the cattle move and graze, they stir up insects from the vegetation, which the egrets then easily catch and eat. The egret benefits, while the cattle are generally unaffected. This is **Commensalism**. This matches with **IV**.

Step 3: Final Answer
The correct matching is:
A matches with I.
B matches with II.
C matches with III.
D matches with IV.
This combination corresponds to **A-I, B-II, C-III, D-IV**, which is option (3).

Quick Tip

When studying ecological interactions, always learn a classic example for each type: Competition (Lion/Leopard), Mutualism (Mycorrhiza, Lichens), Commensalism (Cattle Egret/Cattle, Orchid on a tree), Parasitism (Cuckoo/Crow, Ticks on a dog), Predation (Lion/Deer), Amensalism (Penicillium/Bacteria).

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184. Given below are two statements: one is labelled as Assertion A and the other is labelled as Reason R.
Assertion A: Amniocentesis for sex determination is one of the strategies of Reproductive and Child Health Care Programme.
Reason R: Ban on amniocentesis checks increasing menace of female foeticide.
In the light of the above statements, choose the correct answer from the options given below:

(A) A is true but R is false.
(B) A is false but R is true.
(C) Both A and R are true and R is the correct explanation of A.
(D) Both A and R are true but R is NOT the correct explanation of A.

Correct Answer: (2) A is false but R is true.

Solution:

Step 1: Understanding the Question:
This question asks to evaluate an Assertion and a Reason concerning amniocentesis, its use for sex determination, and its regulation under the Reproductive and Child Health Care (RCH) Programme.

Step 2: Detailed Explanation:

- **Analysis of Assertion (A):** "Amniocentesis for sex determination is one of the strategies of Reproductive and Child Health Care Programme." This statement is false. The RCH Programme aims to create awareness and provide facilities for building a reproductively healthy society. Amniocentesis is a prenatal diagnostic technique used to detect fetal chromosomal abnormalities. Its misuse for sex determination is a crime and leads to female foeticide. Therefore, the RCH Programme actively discourages and works to prevent sex determination via amniocentesis; it is not a "strategy" of the programme.

- **Analysis of Reason (R):** "Ban on amniocentesis checks increasing menace of female foeticide." This statement is true. To prevent the misuse of amniocentesis for sex-selective abortions, the Indian government has imposed a statutory ban on this practice under the Pre-Conception and Pre-Natal Diagnostic Techniques (PCPNDT) Act, 1994. The primary goal of this ban is to curb female foeticide and improve the declining child sex ratio.

Step 3: Final Answer:
Assertion A is false, and Reason R is true. This corresponds to option (2).

> **Quick Tip**
>
> Be very careful with the wording in questions about government programs and policies. The RCH Programme deals with the *topic* of amniocentesis by promoting awareness about its misuse, but it does not include sex determination as a 'strategy'. In fact, its strategy is the exact opposite: to prevent it.

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185. Broad palm with single palm crease is visible in a person suffering from-

(A) Klinefelter's syndrome
(B) Thalassemia
(C) Down's syndrome
(D) Turner's syndrome

Correct Answer: (3) Down's syndrome

Solution:

Step 1: Understanding the Question
The question asks to identify the genetic disorder associated with the physical characteristic of a broad palm with a single, transverse palmar crease.

Step 2: Detailed Explanation
Let's analyze the options:

- **(1) Klinefelter's syndrome (47, XXY):** This affects males and is characterized by features like tall stature, underdeveloped testes, and some feminine characteristics (gynaecomastia). It is not associated with a single palmar crease.

- **(2) Thalassemia:** This is an autosomal recessive blood disorder affecting hemoglobin production. It doesn't have the described palmar characteristic.

- **(3) Down's syndrome (Trisomy 21):** This is caused by the presence of an extra copy of chromosome 21. It is characterized by a set of distinct physical features, including a small round head, furrowed tongue, partially open mouth, and a broad palm with a characteristic single transverse crease (also known as a simian crease).

- **(4) Turner's syndrome (45, XO):** This affects females and is characterized by short stature, a webbed neck, and underdeveloped ovaries. It is not associated with a single palmar crease.

Step 3: Final Answer
The specific physical trait of a broad palm with a single palmar crease is a well-known clinical symptom of Down's syndrome. Therefore, option (3) is the correct answer.

Quick Tip

Associate key physical stigmata with common genetic syndromes. For Down's syndrome, remember "simian crease," flat facial profile, and epicanthal folds. For Turner's syndrome, remember "webbed neck" and short stature.

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186. Select the correct statements with reference to chordates.
A. Presence of a mid-dorsal, solid and double nerve cord.
B. Presence of closed circulatory system.
C. Presence of paired pharyngeal gillslits.
D. Presence of dorsal heart.
E. Triploblastic pseudocoelomate animals.
Choose the correct answer from the options given below:

(A) B, D and E only
(B) C, D and E only
(C) A, C and D only
(D) B and C only

Correct Answer: (4) B and C only

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statements that describe the characteristics of the phylum Chordata.

Step 2: Detailed Explanation:
Let's analyze each statement:

- **A. Presence of a mid-dorsal, solid and double nerve cord.** This is incorrect. The characteristic nerve cord of chordates is **dorsal, hollow, and single**. A solid, ventral, double nerve cord is characteristic of many non-chordates (like annelids and arthropods).

- **B. Presence of closed circulatory system.** This is correct. Vertebrates, a subphylum of Chordata, all have a closed circulatory system where blood is confined to vessels.

- **C. Presence of paired pharyngeal gill slits.** This is correct. This is one of the three fundamental diagnostic characters of chordates, present at some stage in the life of all chordates. In terrestrial vertebrates, they are present only in the embryonic stage.

- **D. Presence of dorsal heart.** This is incorrect. Chordates have a **ventral** muscular heart. A dorsal heart is found in many non-chordates like arthropods.

- **E. Triploblastic pseudocoelomate animals.** This is incorrect. Chordates are triploblastic (have three germ layers), but they are **eucoelomate** (possess a true coelom). Pseudocoelomates include phyla like Aschelminthes.

The only correct statements are B and C.

Step 3: Final Answer:
The correct option is (4), which includes statements B and C only.

Quick Tip

Memorize the three fundamental characteristics of chordates: 1) Presence of a notochord, 2) A dorsal hollow nerve cord, and 3) Paired pharyngeal gill slits. Also, remember the key differences between chordates and non-chordates (e.g., ventral vs. dorsal heart, nerve cord properties).

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187. Which one of the following is the sequence on corresponding coding strand, if the sequence on mRNA formed is as follows 5' AUCGAUCGAUCGAUCGAUCGAUCG AUCG AUCG 3'?

(A) 5' ATCGATCGATCGATCGATCG ATCGATCG 3'
(B) 3' ATCGATCGATCGATCGATCG ATCGATCG 5'
(C) 5' UAGCUAGCUAGCUAGCUAGCUA GCUAGC UAGC 3'
(D) 3' UAGCUAGCUAGCUAGCUAGCUA GCUAGCUAGC 5'

Correct Answer: (1) 5' ATCGATCGATCGATCGATCGATCGATCG ATCGATCG 3'

Solution:

Step 1: Understanding the Question:
The question provides an mRNA sequence and asks for the sequence of the corresponding coding strand of the DNA from which it was transcribed.

Step 2: Key Formula or Approach:
During transcription, the mRNA is synthesized complementary to the template (non-coding) DNA strand. This means the coding (non-template) DNA strand has a sequence that is identical to the mRNA sequence, with two key differences:

1. The sugar is deoxyribose instead of ribose.
2. The nitrogenous base Thymine (T) is present instead of Uracil (U).

The polarity (5' to 3' direction) of the coding strand is the same as the mRNA.

Step 3: Detailed Explanation:
Given mRNA sequence: 5' AUCG... 3'
To find the coding strand sequence, we simply replace every 'U' with a 'T' and keep the polarity the same.
mRNA: **5' AUCGAUCGAUCGAUCGAUCGAUCG AUCG AUCG 3'**
Coding Strand: **5' ATCGATCGATCGATCGATCGATCG ATCG ATCG 3'**
(Note: There seems to be a slight mismatch in the number of repeats between the question stem and the options, but the principle remains the same).
Let's check the options:

- (1) 5' ATCGATCGATCGATCGATCG ATCGATCG 3': This sequence correctly replaces U with T and maintains the 5' to 3' polarity.
- (2) 3' ATCG... 5': The polarity is incorrect.
- (3) and (4): These options contain Uracil (U), which is not found in DNA.

Step 4: Final Answer:
Based on the rule, option (1) represents the correct sequence and polarity for the corresponding coding strand.

Quick Tip

Remember this simple rule: Coding Strand is the "same as mRNA, but with T instead of U". The template strand is complementary to the mRNA.

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188. The parts of human brain that helps in regulation of sexual behaviour, expression of excitement, pleasure, rage, fear etc. are :

(A) Brain stem & epithalamus
(B) Corpus callosum and thalamus
(C) Limbic system & hypothalamus
(D) Corpora quadrigemina & hippocampus

Correct Answer: (3) Limbic system hypothalamus

Solution:

Step 1: Understanding the Question:
The question asks to identify the parts of the human brain responsible for controlling emotions (excitement, pleasure, rage, fear) and sexual behavior.

Step 2: Detailed Explanation:
Let's analyze the functions of the brain parts in the given options:

- **(1) Brain stem & epithalamus:** The brain stem controls vital autonomic functions like breathing and heart rate. The epithalamus is involved in the sleep-wake cycle. They are not the primary centers for emotion.

- **(2) Corpus callosum and thalamus:** The corpus callosum connects the cerebral hemispheres. The thalamus is a major relay station for sensory information.

- **(3) Limbic system & hypothalamus:** This is the correct answer. The limbic system, which includes structures like the amygdala and hippocampus, is often called the "emotional brain." It is responsible for generating emotional responses. The hypothalamus, which is closely associated with the limbic system, regulates many homeostatic functions and basic drives, including sexual behavior and the expression of emotions like rage and pleasure.

- **(4) Corpora quadrigemina & hippocampus:** The corpora quadrigemina in the midbrain are reflex centers for vision and hearing. While the hippocampus is part of the limbic system and is crucial for memory, this option as a whole is less accurate and complete than option (3).

Step 3: Final Answer:
The limbic system and hypothalamus together are the primary centers for regulating emotions and sexual behavior. Therefore, option (3) is correct.

> **Quick Tip**
>
> Associate key terms with brain functions: Brain Stem = Vital functions. Cerebellum = Balance. Cerebrum = Thought/Consciousness. Limbic System + Hypothalamus = Emotions and Drives.

---

189. Which of the following statements are correct?
A. An excessive loss of body fluid from the body switches off osmoreceptors.
B. ADH facilitates water reabsorption to prevent diuresis.
C. ANF causes vasodilation.

D. ADH causes increase in blood pressure.
E. ADH is responsible for decrease in GFR.
Choose the correct answer from the options given below:

(A) A, B and E only
(B) C, D and E only
(C) A and B only
(D) B, C and D only

Correct Answer: (4) B, C and D only

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statements related to the hormonal regulation of water balance and blood pressure.

Step 2: Detailed Explanation:
Let's evaluate each statement:

- **A. An excessive loss of body fluid from the body switches off osmoreceptors.** This is **incorrect**. Excessive fluid loss (dehydration) increases the osmolarity of the blood, which **stimulates** or "switches on" osmoreceptors in the hypothalamus, leading to the release of ADH.

- **B. ADH facilitates water reabsorption to prevent diuresis.** This is **correct**. Antidiuretic Hormone (ADH) increases the permeability of the distal convoluted tubules and collecting ducts to water, promoting water reabsorption and reducing urine output (diuresis).

- **C. ANF causes vasodilation.** This is **correct**. Atrial Natriuretic Factor (ANF), released from the heart's atria in response to high blood pressure, causes the dilation of blood vessels (vasodilation), which helps to lower blood pressure.

- **D. ADH causes increase in blood pressure.** This is **correct**. At high concentrations, ADH acts as a vasoconstrictor, increasing peripheral resistance and thereby raising blood pressure. This is why it is also known as vasopressin.

- **E. ADH is responsible for decrease in GFR.** This is **incorrect**. ADH's vasoconstrictor effect generally increases systemic blood pressure, which would tend to maintain or increase the Glomerular Filtration Rate (GFR). Its primary role is not to decrease GFR.

The correct statements are B, C, and D.

Step 3: Final Answer:
The option containing only the correct statements is (4).

> **Quick Tip**
>
> Remember the antagonistic relationship between ADH/RAAS and ANF. ADH and RAAS work to increase blood pressure and conserve water. ANF works to decrease blood pressure and promote water loss.

---

190. Which of the following statements are correct?
A. Basophils are most abundant cells of the total WBCs
B. Basophils secrete histamine, serotonin and heparin
C. Basophils are involved in inflammatory response
D. Basophils have kidney shaped nucleus
E. Basophils are agranulocytes
Choose the correct answer from the options given below:

(A) B and C only
(B) A and B only
(C) D and E only
(D) C and E only

Correct Answer: (1) B and C only

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statements about basophils, a type of white blood cell (WBC).

Step 2: Detailed Explanation:
Let's evaluate each statement:

- **A. Basophils are most abundant cells of the total WBCs.** This is **incorrect**. Neutrophils are the most abundant WBCs (60-65%), while basophils are the least abundant (0.5-1%).

- **B. Basophils secrete histamine, serotonin and heparin.** This is **correct**. These chemicals are stored in the granules of basophils.

- **C. Basophils are involved in inflammatory response.** This is **correct**. By releasing histamine, a vasodilator, they mediate inflammatory reactions.

- **D. Basophils have kidney shaped nucleus.** This is **incorrect**. Monocytes have a large, kidney-shaped nucleus. The nucleus of a basophil is typically S-shaped or bilobed, but it is often obscured by its large, coarse granules.

- **E. Basophils are agranulocytes.** This is **incorrect**. Basophils are classified as granulocytes because of the presence of prominent granules in their cytoplasm, along with neutrophils and eosinophils.

Therefore, only statements B and C are correct.

Step 3: Final Answer:
The combination of correct statements is B and C, which corresponds to option (1).

> Quick Tip
>
> Use the mnemonic "Never Let Monkeys Eat Bananas" to remember the decreasing order of abundance of WBCs: Neutrophils > Lymphocytes > Monocytes > Eosinophils > Basophils.

---

191. Which of the following are NOT under the control of thyroid hormone?
A. Maintenance of water and electrolyte balance
B. Regulation of basal metabolic rate
C. Normal rhythm of sleep-wake cycle
D. Development of immune system
E. Support the process of R.B.Cs formation
Choose the correct answer from the options given below:

(A) C and D only
(B) D and E only
(C) A and D only
(D) B and C only

Correct Answer: (1) C and D only

Solution:

Step 1: Understanding the Question:
The question asks to identify which of the listed physiological processes are NOT regulated by hormones from the thyroid gland.

Step 2: Detailed Explanation:
Let's analyze the role of thyroid hormones in each process:

- **A. Maintenance of water and electrolyte balance:** Thyroid hormones do influence this process, though the primary control is by ADH and aldosterone. This is considered a function.

- **B. Regulation of basal metabolic rate (BMR):** This is a **primary and major function** of thyroid hormones (thyroxine and triiodothyronine).

- **C. Normal rhythm of sleep-wake cycle:** This circadian rhythm is primarily regulated by the hormone **melatonin** from the pineal gland, not by thyroid hormones.

- **D. Development of immune system:** The development and maturation of T-lymphocytes, a key part of the immune system, is regulated by hormones like **thymosin** from the thymus gland, not by thyroid hormones.

- **E. Support the process of R.B.Cs formation:** Thyroid hormones are known to **support erythropoiesis** (RBC formation).

Based on this analysis, the sleep-wake cycle (C) and the development of the immune system (D) are not under the control of thyroid hormones.

Step 3: Final Answer:
The correct option that lists functions not controlled by the thyroid hormone is (1), which includes C and D only.

> Quick Tip
>
> When studying hormones, create a table listing the gland, the hormone(s) it secretes, and the primary functions of each hormone. This helps in quickly differentiating the roles, e.g., Thyroid $\rightarrow$ BMR, Pineal $\rightarrow$ Sleep, Thymus $\rightarrow$ Immunity.

---

192. Which of the following is characteristic feature of cockroach regarding sexual dimorphism ?

(A) Presence of sclerites
(B) Presence of anal cerci
(C) Dark brown body colour and anal cerci
(D) Presence of anal styles

Correct Answer: (4) Presence of anal styles

Solution:

Step 1: Understanding the Question:
The question asks to identify a feature that distinguishes male and female cockroaches (sexual dimorphism).

Step 2: Detailed Explanation:
Let's examine the features listed:

- **(1) Presence of sclerites:** Sclerites are the hardened plates of the exoskeleton. They are present in both male and female cockroaches.
- **(2) Presence of anal cerci:** Anal cerci are a pair of jointed filamentous structures that arise from the 10th abdominal tergum. They are sensory in function and are present in **both** sexes.
- **(3) Dark brown body colour and anal cerci:** Body colour can vary and is not a reliable dimorphic feature. As mentioned, anal cerci are present in both sexes.
- **(4) Presence of anal styles:** Anal styles are a pair of short, thread-like, unjointed structures found on the 9th abdominal sternum. They are present **only in male** cockroaches. Their absence in females makes them a key feature of sexual dimorphism.

Step 3: Final Answer:
The presence of anal styles is a characteristic feature of male cockroaches only. Therefore, it is a feature of sexual dimorphism, and option (4) is the correct answer.

Quick Tip

A simple way to remember the key difference in the posterior abdomen of a cockroach: Anal Cerci are present in both sexes (C for common/couple), while Anal Styles are present only in males (S for single/solo sex).

---

193. Which one of the following is NOT an advantage of inbreeding?

(A) Elimination of less desirable genes and accumulation of superior genes takes place due to it.
(B) It decreases the productivity of inbred population, after continuous inbreeding.
(C) It decreases homozygosity.
(D) It exposes harmful recessive genes that are eliminated by selection.

Correct Answer: (2) It decreases the productivity of inbred population, after continuous inbreeding.

Solution:

Step 1: Understanding the Question:
The question asks to identify the statement that describes a disadvantage (or something that is not an advantage) of inbreeding. Inbreeding refers to the mating of more closely related individuals within the same breed for several generations.

Step 2: Detailed Explanation:
Let's analyze the effects described in each option:

- **(1) Elimination of less desirable genes and accumulation of superior genes takes place due to it.** This is a primary **advantage** of inbreeding. It increases homozygosity, which helps in creating pure lines and accumulating desirable genes.

- **(2) It decreases the productivity of inbred population, after continuous inbreeding.** This phenomenon is known as **inbreeding depression**. It is a major **disadvantage** of continuous inbreeding, leading to reduced fertility and productivity. Since it's a disadvantage, it is NOT an advantage.

- **(3) It decreases homozygosity.** This statement is factually incorrect. Inbreeding **increases** homozygosity and decreases heterozygosity.

- **(4) It exposes harmful recessive genes that are eliminated by selection.** This is an **advantage**. Because inbreeding increases homozygosity, harmful recessive alleles are expressed in the phenotype. This allows breeders to identify and eliminate these undesirable alleles from the population.

The question asks for what is NOT an advantage. Option (2) describes a clear disadvantage.

Step 3: Final Answer:
The decrease in productivity due to inbreeding depression is a disadvantage, not an advantage. Therefore, option (2) is the correct answer.

> **Quick Tip**
>
> Remember the pros and cons of inbreeding. Pros: increases homozygosity, creates pure lines, exposes harmful recessive alleles for selection. Con: Inbreeding depression (reduced fertility and productivity).

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194. The unique mammalian characteristics are:

(A) hairs, pinna and indirect development
(B) pinna, monocondylic skull and mammary glands
(C) hairs, tympanic membrane and mammary glands
(D) hairs, pinna and mammary glands

Correct Answer: (4) hairs, pinna and mammary glands

Solution:

Step 1: Understanding the Question:
The question asks to identify a set of characteristics that are unique or defining features of the class Mammalia.

Step 2: Detailed Explanation:
Let's analyze the characteristics in each option:

- **(1) hairs, pinna and indirect development:** Most mammals exhibit **direct development** (no larval stage), so this is incorrect.

- **(2) pinna, monocondylic skull and mammary glands:** Mammals have a **dicondylic skull** (two occipital condyles that articulate with the vertebral column). A monocondylic skull is found in reptiles and birds. This is incorrect.

- **(3) hairs, tympanic membrane and mammary glands:** The **tympanic membrane** (eardrum) is present in many other vertebrates like frogs, reptiles, and birds. It is not a unique characteristic of mammals. This is incorrect.

- **(4) hairs, pinna and mammary glands:**
  - **Hairs:** The presence of hair or fur on the body is a unique and defining feature of mammals.
  - **Pinna:** The presence of external ears (pinnae) is characteristic of most mammals.
  - **Mammary glands:** The presence of milk-producing mammary glands is the most fundamental and unique characteristic of mammals, giving the class its name.

  This set contains defining characteristics of mammals.

Step 3: Final Answer:
The combination of hairs, pinna, and mammary glands represents the unique characteristics of mammals among the given choices. Therefore, option (4) is correct.

Quick Tip

The two most absolute defining features of mammals are the presence of hair and mammary glands. Other key features to remember include a four-chambered heart, a diaphragm for breathing, and a dicondylic skull.

---

195. Select the correct statements.
A. Tetrad formation is seen during Leptotene.
B. During Anaphase, the centromeres split and chromatids separate.
C. Terminalization takes place during Pachytene.
D. Nucleolus, Golgi complex and ER are reformed during Telophase.
E. Crossing over takes place between sister chromatids of homologous chromosome.
Choose the correct answer from the options given below:

(A) A, C and E only
(B) B and E only
(C) A and C only
(D) B and D only

Correct Answer: (4) B and D only

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statements about the events of meiosis from a given list.

Step 2: Detailed Explanation:
Let's evaluate each statement:

- **A. Tetrad formation is seen during Leptotene.** This is incorrect. Tetrads, or bivalents, are formed when homologous chromosomes pair up (synapsis). This process occurs during the **Zygotene** stage of Prophase I.

- **B. During Anaphase, the centromeres split and chromatids separate.** This statement is correct. This event specifically occurs during **Anaphase II** of meiosis (and also Anaphase of mitosis). In Anaphase I, homologous chromosomes separate, but sister chromatids remain joined at the centromere. Since this event does happen in one of the anaphase stages of meiosis, the statement is considered correct.

- **C. Terminalization takes place during Pachytene.** This is incorrect. Crossing over occurs during Pachytene. The movement of chiasmata towards the end of the chromosomes (terminalization) occurs during **Diplotene** and is completed by **Diakinesis**.

- **D. Nucleolus, Golgi complex and ER are reformed during Telophase.** This is correct. At the end of both Meiosis I (in Telophase I) and Meiosis II (in Telophase II), the nuclear envelope and nucleolus reappear, and the cell organelles are reformed, followed by cytokinesis.

- **E. Crossing over takes place between sister chromatids of homologous chromosome.** This is incorrect. Crossing over is the exchange of genetic material between **non-sister chromatids** of homologous chromosomes. This is a crucial source of genetic recombination.

The correct statements are B and D.

Step 3: Final Answer:
The correct option containing only the correct statements is (4) B and D only.

> **Quick Tip**
>
> Remember the sequence and key events of Prophase I: Leptotene (condensation), Zygotene (synapsis), Pachytene (crossing over), Diplotene (chiasmata visible), Diakinesis (terminalization). Also, critically distinguish between Meiosis I (separation of homologous chromosomes) and Meiosis II (separation of sister chromatids).

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196. Given below are two statements:
Statement I: During G_0 phase of cell cycle, the cell is metabolically inactive.
Statement II: The centrosome undergoes duplication during S phase of interphase.
In the light of the above statements, choose the most appropriate answer from the options given below:

(A) Statement I is correct but Statement II is incorrect.
(B) Statement I is incorrect but Statement II is correct.
(C) Both Statement I and Statement II are correct.
(D) Both Statement I and Statement II are incorrect.

Correct Answer: (2) Statement I is incorrect but Statement II is correct.

Solution:

Step 1: Understanding the Question:
The question asks to evaluate the correctness of two statements regarding events in the cell cycle.

Step 2: Detailed Explanation:

- **Analysis of Statement I:** The G_0 phase, or quiescent stage, is a non-dividing state that cells enter from the G_1 phase. Cells in this phase exit the cell cycle and do not proliferate. However, they are not metabolically inactive. In fact, they are metabolically active and perform their specialized physiological functions (e.g., a neuron conducts nerve impulses, a liver cell performs metabolic conversions). Therefore, Statement I is incorrect.

- **Analysis of Statement II:** The S phase (Synthesis phase) of interphase is primarily known for DNA replication. Along with DNA replication, the centrosome also duplicates during the S phase in the cytoplasm. The two centrosomes then move to opposite poles of the cell to form the mitotic spindle during the M phase. Therefore, Statement II is correct.

Step 3: Final Answer:
Since Statement I is incorrect and Statement II is correct, option (2) is the correct answer.

Quick Tip

It is crucial to distinguish between being 'proliferatively inactive' and 'metabolically inactive'. G_0 cells are the former, not the latter. They are workhorse cells of the body, just not dividing.

---

197. Which of the following statements are correct regarding skeletal muscle?
A. Muscle bundles are held together by collagenous connective tissue layer called fascicle.
B. Sarcoplasmic reticulum of muscle fibre is a store house of calcium ions.
C. Striated appearance of skeletal muscle fibre is due to distribution pattern of actin and myosin proteins.
D. M line is considered as functional unit of contraction called sarcomere.
Choose the most appropriate answer from the options given below:

(A) A, C and D only
(B) C and D only
(C) A, B and C only
(D) B and C only

Correct Answer: (4) B and C only

Solution:

Step 1: Understanding the Question:
The question asks to identify the correct statements about the structure and function of skeletal muscle.

Step 2: Detailed Explanation:
Let's evaluate each statement:

- **A. Muscle bundles are held together by collagenous connective tissue layer called fascicle.** This is **incorrect**. The muscle bundles themselves are called **fascicles**. The connective tissue layer that holds them together is the **perimysium**.

- **B. Sarcoplasmic reticulum of muscle fibre is a store house of calcium ions.** This is **correct**. The sarcoplasmic reticulum sequesters and stores calcium ions (Ca^{2+}). Their release into the sarcoplasm initiates muscle contraction.

- **C. Striated appearance of skeletal muscle fibre is due to distribution pattern of actin and myosin proteins.** This is **correct**. The alternating light (I) bands (containing thin actin filaments) and dark (A) bands (containing thick myosin filaments and overlapping actin) create the characteristic striated pattern.

- **D. M line is considered as functional unit of contraction called sarcomere.** This is **incorrect**. The functional unit of contraction is the **sarcomere**, which is defined as the region of a myofibril between two successive **Z lines**. The M line is a protein structure in the middle of the H zone that holds thick filaments together.

Therefore, only statements B and C are correct.

Step 3: Final Answer:
The correct option is (4), which includes statements B and C only.

Quick Tip

For muscle structure, remember the hierarchy: Muscle → Fascicle (bundle of fibers) → Muscle Fiber (cell) → Myofibril → Sarcomere. The sarcomere is the key functional unit, running from Z-line to Z-line.

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198. Match List I with List II.

	List I		**List II**
A.	**Logistic growth**	**I.**	**Unlimited resource availability condition**
B.	**Exponential growth**	**II.**	**Limited resource availability condition**
C.	**Expanding age pyramid**	**III.**	**The percent individuals of pre-reproductive age is largest**
D.	**Stable age pyramid**	**IV.**	**The percent individuals of pre-reproductives and reproductive... are s**

Choose the correct answer from the options given below:

(A) A-II, B-IV, C-I, D-III
(B) A-II, B-IV, C-III, D-I

(C) A-II, B-I, C-III, D-IV
(D) A-II, B-III, C-I, D-IV

Correct Answer: (3) A-II, B-I, C-III, D-IV

Solution:

Step 1: Understanding the Question:
The question requires matching concepts from population ecology in List I with their correct descriptions in List II.

Step 2: Detailed Explanation:

- **A. Logistic growth:** This model of population growth is considered more realistic as it assumes that resources are finite. Population growth slows as it approaches the environment's carrying capacity (K). Thus, it corresponds to a **Limited resource availability condition**. So, **A matches with II**.

- **B. Exponential growth:** This model describes population growth in an idealized environment with no limiting factors. It assumes **Unlimited resource availability**. So, **B matches with I**.

- **C. Expanding age pyramid:** This type of pyramid has a very broad base and tapers towards the top. The broad base indicates that the percentage of **pre-reproductive individuals is the largest**, signifying high birth rates and future population growth. So, **C matches with III**.

- **D. Stable age pyramid:** This pyramid has a bell shape, where the number of **pre-reproductive and reproductive individuals are roughly the same**. This indicates that the birth rate is almost equal to the death rate, and the population size is stable. So, **D matches with IV**.

The correct set of matches is A-II, B-I, C-III, D-IV.

Step 3: Final Answer:
This combination corresponds to option (3).

Quick Tip

Remember the shapes of population models: Exponential growth = J-shaped curve (unlimited resources). Logistic growth = S-shaped curve (limited resources). Expanding pyramid = triangle shape. Stable pyramid = bell shape. Declining pyramid = urn shape.

199. Match List I with List II.

List I		List II	
A.	**Mast cells**	**I.**	**Ciliated epithelium**
B.	**Inner surface of bronchiole**	**II.**	**Areolar connective tissue**
C.	**Blood**	**III.**	**Cuboidal epithelium**
D.	**Tubular parts of nephron**	**IV.**	**specialised connective tissue**

Choose the correct answer from the options give below:

(A) A-II, B-I, C-IV, D-III
(B) A-III, B-IV, C-II, D-I
(C) A-I, B-II, C-IV, D-III
(D) A-II, B-III, C-I, D-IV

Correct Answer: (1) A-II, B-I, C-IV, D-III

Solution:

Step 1: Understanding the Question:
The question requires matching the structures or cell types in List I with the corresponding type of animal tissue in List II.

Step 2: Detailed Explanation:
Let's analyze each item to find the correct match:

- **A. Mast cells:** These are specialized cells that secrete histamine, heparin, and serotonin. They are found in connective tissue, specifically in Areolar connective tissue. So, **A matches with II**.

- **B. Inner surface of bronchiole:** The smaller bronchioles are lined with Ciliated epithelium. The cilia help in moving mucus and trapped particles out of the respiratory tract. So, **B matches with I**.

- **C. Blood:** Blood is a fluid connective tissue consisting of plasma, red blood cells, white blood cells, and platelets. It is classified as a specialised connective tissue. So, **C matches with IV**.

- **D. Tubular parts of nephron:** The tubular parts of the nephron in the kidney, such as the Proximal Convoluted Tubule (PCT) and Distal Convoluted Tubule (DCT), are primarily lined with Cuboidal epithelium, which is involved in secretion and absorption. So, **D matches with III**.

Combining the matches, we get: A-II, B-I, C-IV, D-III.

Step 3: Final Answer:
The correct combination is A-II, B-I, C-IV, D-III, which corresponds to option (1).

> **Quick Tip**
>
> For tissue-related questions, create a chart linking different types of epithelial, connective, muscular, and neural tissues to their specific locations and functions in the body. This is a high-yield topic for matching questions.

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200. In cockroach, excretion is brought about by-
A. Phallic gland
B. Urecose gland
C. Nephrocytes
D. Fat body
E. Collaterial glands
Choose the correct answer from the options given below:

(A) B, C and D only
(B) B and D only
(C) A and E only
(D) A, B and E only

Correct Answer: (1) B, C and D only

Solution:

Step 1: Understanding the Question:
The question asks to identify the structures involved in excretion in cockroaches from the given list.

Step 2: Detailed Explanation:
Let's examine the function of each structure listed:

- **A. Phallic gland:** This is an accessory reproductive gland in male cockroaches, contributing to the formation of the spermatophore. It has a reproductive, not excretory, function.

- **B. Urecose gland:** Located in the mushroom-shaped gland of male cockroaches, these glands store uric acid and release it with the spermatophore. They are considered accessory excretory organs.

- **C. Nephrocytes:** These are cells found in the body cavity that absorb nitrogenous wastes from the haemolymph and store them until they can be further processed or deposited.

They play a role in excretion.

- **D. Fat body:** The fat body in cockroaches has multiple functions, including metabolism and storage of nutrients. It also contains urate cells that synthesize and store uric acid, thus playing a significant role in excretion.

- **E. Collaterial glands:** These are accessory reproductive glands in female cockroaches. Their secretion forms the ootheca (egg case). They have a reproductive, not excretory, function.

The primary excretory organs in cockroaches are the Malpighian tubules (not listed). However, among the given options, the Urecose gland, Nephrocytes, and Fat body are all involved in excretion.

Step 3: Final Answer:
The structures involved in excretion are B, C, and D. Therefore, option (1) is the correct answer.

Quick Tip

When asked about excretion in insects like cockroaches, remember the main organs are Malpighian tubules. However, be aware of the accessory excretory structures: Fat body, Nephrocytes, and Urecose glands. Differentiate these from reproductive glands like the Phallic gland and Collaterial glands.
