Image Generation using Stable Diffusion (Text to Image)

Objective:

Generate realistic images from a text prompt using Hugging Face's Stable Diffusion model, and deploy with an elegant Gradio interface.

Environment Setup:

Use Google Colab with GPU runtime, as this model is too large for CPU.

```
!pip install transformers diffusers accelerate gradio
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from diffusers import StableDiffusionPipeline
import torch
import gradio as gr
# Load the Model (Use public model)
model_id = "runwayml/stable-diffusion-v1-5"
pipe = StableDiffusionPipeline.from_pretrained(
    model_id,
    torch_dtype=torch.float16
).to("cuda")
     Loading pipeline components...: 100%
                                                                              7/7 [00:14<00:00, 2.38s/it]
# Define Image Generation Function
def generate_image(prompt):
    image = pipe(prompt).images[0]
    return image
gr.Interface(
    fn=generate_image,
    inputs=gr.Textbox(
        placeholder="Describe an image (e.g., A panda astronaut on Mars)",
        label="| Enter your Prompt"
    ),
    outputs="image",
    title="{} AI Image Generator",
    description="Type a creative prompt and get an AI-generated image using Stable Diffusion.",
).launch()
```



🚁 It looks like you are running Gradio on a hosted Jupyter notebook, which requires `share=True`. Automatically setting `share=True` (you

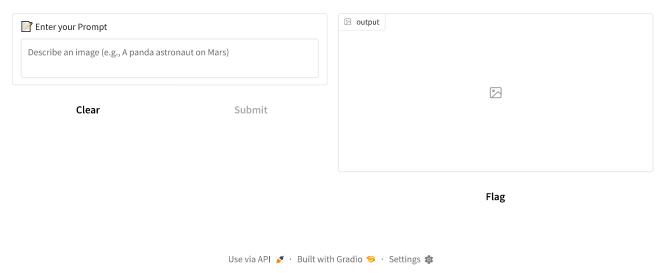
Colab notebook detected. To show errors in colab notebook, set debug=True in launch()

* Running on public URL: https://764c16509eff254bcf.gradio.live

This share link expires in 1 week. For free permanent hosting and GPU upgrades, run `gradio deploy` from the terminal in the working dir

Al Image Generator

Type a creative prompt and get an Al-generated image using Stable Diffusion.



Start coding or generate with AI.