

# Stable Diffusion

Stable Diffusion is an AI model for text to image generation. There are various GUIs available to tweak settings, add specific styles like Disney or photorealistic (Lora models) and help adjusting existing images via picture to picture or selection infill.

The guides below focus on the [Automatic1111 Stable Diffusion WebUI](#).

## Linux Docker install

This docker requires docker as well as the GPU drivers installed on the host system.

<https://github.com/AbdBarho/stable-diffusion-webui-docker>

```
git clone https://github.com/AbdBarho/stable-diffusion-webui-docker
cd stable-diffusion-webui-docker
vi docker-compose.yaml
#adjust name to stable-diffusion if desired
#adjust data and output dir location to /opt/stable-diffusion/data and
output respectively if desired
mkdir -p /opt/stable-diffusion/{data,output}

#Build the ai and download the models.
#Note, first start will take 15-60 minutes to download models to data folder
as cache depending on internet connection
#size approximately 10GB

docker compose --profile download up --build

# if download errors occur, just repeat the command.

# build the desired interface:
# docker compose --profile [ui] up --build
# where [ui] is one of: invoke | auto | auto-cpu | comfy | comfy-cpu

# use auto-cpu or comfy-cpu for cpu only interface

#docker compose --profile auto-cpu up --build
docker compose --profile auto up --build

# later:
docker compose --profile auto up -d
```

access the app on:

<http://localhost:7860/>

# Windows install for AMD GPU

Windows AMD AUTOMATIC1111 stable diffusion webui using Microsoft DirectML for GPU acceleration:

1. Download and run installers for [Python 3.10.6](#) (ticking Add to PATH) and [git](#)
2. Clone the fork as AMD is not officially supported  
Paste into command prompt in a directory where it should live

```
git clone
https://github.com/lshqqytiger/stable-diffusion-webui-directml
cd stable-diffusion-webui-directml
git submodule init
git submodule update
venv\Scripts\python.exe -m pip install --upgrade pip
venv\Scripts\pip.exe pip install -r requirements.txt
venv\Scripts\pip.exe install torch-directml
```

3. Right-click webui-user.bat and adjust the command line args to

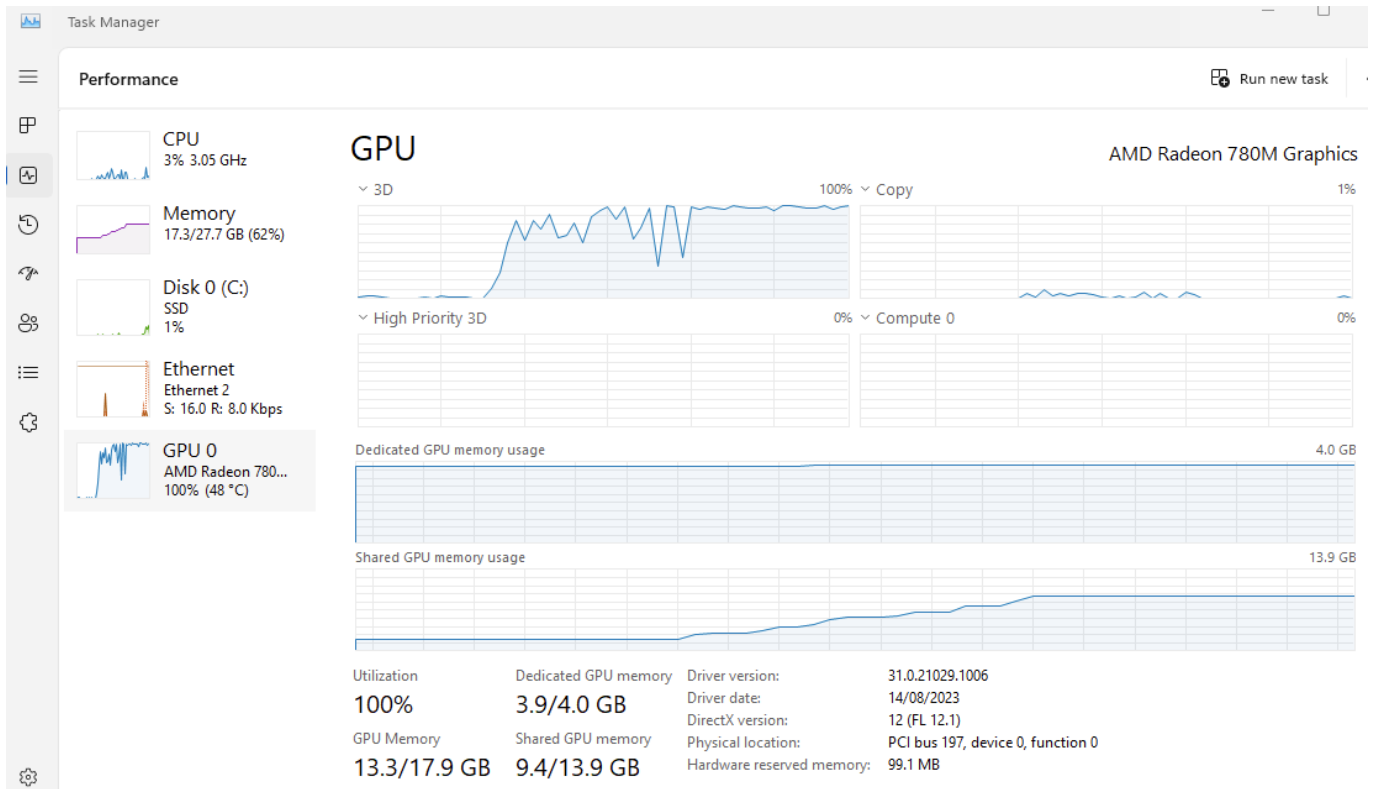
```
COMMANDLINE_ARGS=--use-directml --skip-torch-cuda-test
```

4. If you have only 4-6gb vram, try adding these additional flags to webui-user.bat's command line args like so:

```
--opt-sub-quad-attention --lowvram --disable-nan-check
```

5. Double-click webui-user.bat in that directory, this will download ~2.7GB torch and other python modules, followed by the 4GB Stable-Diffusion 1.5 pruned model.
6. If it looks like it is stuck when installing or running, press enter in the terminal and it should continue.
7. once done, a browser opens to <http://localhost:7860/>

Check/Test if GPU is used in Task Manager's GPU tab. Dedicated GPU Memory should show large usage for the loaded Model and when generating a test image in the web interface, the GPU usage should ramp up like here:



## Performance

Approx times to generate 512x512 default settings images:

CPU/GPU Model	Time
CPU Intel 8th gen i5 (Intel NUC)	5-15 minutes
GPU AMD Radeon 780M	1.5 minutes
GPU AMD Radeon 7900XTX	21.7 seconds
GPU Nvidia RTX 4080	1.2 seconds

## Gallery Addon

As an extension for SD-webui: Open the Extensions tab in SD-webui.

1. Select the Install from URL option.
2. Enter <https://github.com/zanllp/sd-webui-infinite-image-browsing>
3. Click on the Install button.
4. Wait for the installation to complete and click on Apply and restart UI.

## general advice

[Some general Stable Diffusion prompt advice](#)

[Inpainting explanation](#)

# Models / Settings

LoRA models/styles/tools as well as base models can be searched and downloaded from [CivitAI](#)

## Realistic Vision + Lora

Description for [photorealistic images](#):

1. Download model [Realistic Vision V6.0 B1 Base Model](#) from <https://civitai.com/models/4201/realistic-vision-v11> and save it into the models/stable-diffusion directory
2. create models/Lora directory
3. download [LORA Style hyperrealism art](#) and save to models/Lora directory
4. download [LORA epi\\_noiseoffset](#) and save to models/Lora directory
5. in web interface go to extensions → install from url and paste to install:

```
https://github.com/mcmonkeyprojects/sd-dynamic-thresholding  
https://github.com/opparco/stable-diffusion-webui-composable-lora
```

6. then go to installed tab and click apply and quit
7. restart stable diffusion web ui
8. go to Lora tab, configure entries (which will create json files, then the entries disappear)

Base settings:

```
set sampling steps to 20  
set sampling method to DPM++ SDE Karras  
set width to 768 and height to 512  
set CFG scale to 6  
set seed to -1  
enable Dynamic Thresholding (CFG Scale Fix)  
enable all composable Lora
```

## Stable Diffusion SDXL Model

```
cd /opt/stable-diffusion/data/models/Stable-diffusion  
wget  
https://huggingface.co/stabilityai/stable-diffusion-xl-base-1.0/raw/main/sd_xl_base_1.0_0.9vae.safetensors
```

## Stable Diffusion XL Turbo model

This is for general objects/scenes/animals/etc, faster than SDXL

download [SDXL Turbo](https://huggingface.co/stabilityai/sd-xl-turbo) from <https://huggingface.co/stabilityai/sd-xl-turbo> and copy into Models/Stable-diffusion directory.

## Image Style/Prompt libraries

[Lexica PromptHero](#)

### Some test image prompts

```
3d, 8k, masterpiece, (best quality), Japanese garden, waterfall, cherry blossoms in full bloom, conifer, koi pond, (pagoda:1.1), stone lantern, water basin, water reflections, winding path, (night sky:1.:1), hokusai inspiration, ultra-realistic, pastel color scheme, soft lighting, golden hour, tranquil atmosphere, landscape orientation
```

```
EasyNegative, (worst quality:1.2), (low quality:1.2), (lowres:1.1), (monochrome:1.1), (greyscale), multiple views, comic, sketch, watermark
```

```
25 steps, 2m karras, 1024x768 seed 1192013237
```

```
greg rutkowski, highly detailed, dark, surreal scary swamp, terrifying, horror, poorly lit, trending on artstation, incredible composition, masterpiece
```

with settings:

```
LMS Karras  
50 steps  
CFG scale 8  
random seed
```

Test. use in txt2img the following prompt:

```
RAW photo, close up photo of face, woman with dark hair in rain at night, blue eyes, detailed eyes, dark skin, (high detailed skin:1.2), 8k uhd, dslr, soft lighting, high quality, film grain, Fujifilm XT3  
<lora:lora_contrast_fix:1> <lora:lora-style-hyperrealism-art:1>
```

and as negative prompt below:

```
(deformed iris, deformed pupils, semi-realistic, cgi, 3d, render, sketch, cartoon, drawing, anime:1.4), text, close up, cropped, out of frame, worst quality, low quality, jpeg artifacts, ugly, duplicate, morbid, mutilated, extra fingers, mutated hands, poorly drawn hands, poorly drawn face, mutation, deformed, blurry, dehydrated, bad anatomy, bad proportions, extra limbs, cloned face, disfigured, gross proportions, malformed limbs, missing arms, missing legs, extra arms, extra legs, fused fingers, too many fingers, long neck
```

then click generate as often as you like

Stable Diffusion checkpoint  
realistic-vision-6.0b1-base.safetensors [15012c5]

txt2img | img2img | Extras | PNG Info | Checkpoint Merger | Train | Settings | Extensions

50/75  
RAW photo, close up photo of face, woman with dark hair in rain at night, blue eyes, detailed eyes, dark skin, (high detailed skin:1.2), 8k uhd, soft lighting, high quality, film grain, Fujifilm XT3 <!--lora:contrast\_fix:1-->  
<!--lora:style-hyperrealism:art:1-->

128/150  
(deformed iris, deformed pupils, semi-realistic, cgi, 3d, render, sketch, cartoon, drawing, anime:1.4), text, close up, cropped, out of frame, worst quality, low quality, jpeg artifacts, ugly, duplicate, morbid, mutilated, extra fingers, mutated hands, poorly drawn hands, poorly drawn face, mutation, deformed, blurry, dehydrated, bad anatomy, bad proportions, extra limbs, cloned face, disfigured, gross proportions, malformed limbs, missing arms, missing legs, extra arms, extra legs, fused fingers, too many fingers, long neck

Generate

Generation | Textual Inversion | Hypernetworks | Checkpoints | Lora

Sampling method: DPM++ SDE Karras | Sampling steps: 20

Hires. fix |  Refiner


Width: 768 | Height: 512 | Batch count: 1 | Batch size: 1 | CFG Scale: 6

Seed: -1

Dynamic Thresholding (CFG Scale Fix)

Composable Lora

Script: None





00002-859497391.png

551.0 KB ↓

RAW photo, close up photo of face, woman with dark hair in rain at night, blue eyes, detailed eyes, dark skin, (high detailed skin:1.2), 8k uhd, dslr, soft lighting, high quality, film grain, Fujifilm XT3 <lora:lora\_contrast\_fix:1> <lora:lora-style-hyperrealism-art:1>  
Negative prompt: (deformed iris, deformed pupils, semi-realistic, cgi, 3d, render, sketch, cartoon, drawing, anime:1.4), text, close up, cropped, out of frame, worst quality, low quality, jpeg artifacts, ugly, duplicate, morbid, mutilated, extra fingers, mutated hands, poorly drawn hands, poorly drawn face, mutation, deformed, blurry, dehydrated, bad anatomy, bad proportions, extra limbs, cloned face, disfigured, gross proportions, malformed limbs, missing arms, missing legs, extra arms, extra legs, fused fingers, too many fingers, long neck  
Steps: 20, Sampler: DPM++ SDE Karras, CFG scale: 6, Seed: 859497391, Size: 768x512, Model hash: cc6cb27103, Model: v1-5-pruned-emaonly, Lora hashes: "lora\_contrast\_fix: d1131f7207d6, lora-style-hyperrealism-art: 6d4ea812a2d2", Dynamic thresholding enabled: True, Mimic scale: 7, Separate Feature Channels: True, Scaling Startpoint: MEAN, Variability Measure: AD, Interpolate Phi: 1, Threshold percentile: 100, Version: v1.7.0

From:  
<http://wuff.dyndns.org/> - **Wulf's Various Things**

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