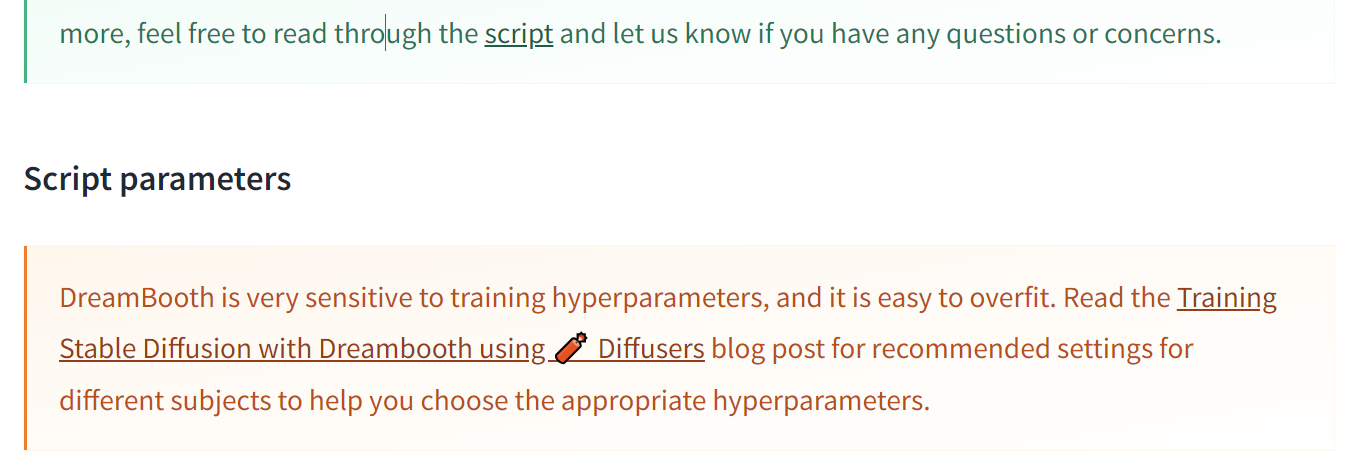
Step 1: Set up environment

1. <https://huggingface.co/docs/diffusers/en/training/dreambooth>
2. Follow the instructions to clone the repo and install the requirement until



Step 2: Fine Tuned model

1. Enter the folder /diffusers/examples/dreambooth
2. Create a script for example (train.sh)

exec >logfile.txt 2>&1 # Log the script for debugging

export MODEL\_NAME='stabilityai/stable-diffusion-2-1'

export INSTANCE\_DIR="./input\_imageset" # The folder for training image dataset

export CLASS\_DIR="./class\_glass" # The folder that store Class image for Prior preservation loss

export OUTPUT\_DIR="./model\_2\_glasses\_of\_wine" # The folder that store the fine tuned model

accelerate launch train\_dreambooth\_lora.py \

--pretrained\_model\_name\_or\_path=$MODEL\_NAME \

--instance\_data\_dir=$INSTANCE\_DIR \

--class\_data\_dir=$CLASS\_DIR \

--output\_dir=$OUTPUT\_DIR \

--with\_prior\_preservation --prior\_loss\_weight=1.0 \

--instance\_prompt="2 glasses of wine" \

--class\_prompt="glasses of wine" \

--resolution=768 \

--train\_batch\_size=1 \

--gradient\_accumulation\_steps=2 --gradient\_checkpointing \

--use\_8bit\_adam \

--learning\_rate=5e-6 \

--lr\_scheduler="constant" \

--lr\_warmup\_steps=0 \

--num\_class\_images=20 \

--max\_train\_steps=50\

--checkpointing\_steps=25 \

1. The learning\_rate and max\_train\_steps affect the result significantly
2. Change the instance\_prompt and class\_prompt accordingly to your task

Step 3: Inference

1. Write a python script for inference (inference.py)
2. Here is a example

from diffusers import StableDiffusionPipeline

import os

model\_id = "stabilityai/stable-diffusion-2-1"

pipe = StableDiffusionPipeline.from\_pretrained(model\_id).to("cuda")

pipe.load\_lora\_weights("./fined\_tuned\_model/checkpoint-50") # Choose a checkpoint

prompt = "2 glasses of wine"

prefix = "fine\_tuned\_2\_glasses\_of\_wine"

if not os.path.exists(prefix):

os.mkdir(prefix)

for i in range(10):

pipe(prompt, num\_inference\_steps=50, guidance\_scale=7.5).images[0].save(prefix+"/"+str(i)+".png")

1. Check the generated image with the original image
   1. You can generate image using original diffusion model (ie. stable-diffusion-2-1) by commenting out this pipe.load\_lora\_weights("./fined\_tuned\_model/checkpoint-")