

Department of Computer Science



Submitted in part fulfilment for the degree of
MSc in Software Engineering.

Predicting Memorable Regions Of Images Using Deep Learning and Adversarial Networks

Babar Khan

Version 1, 2023-April-11

Supervisor: Adrian Bors

Contents

1	Introduction	iv
1.1	Background	iv
1.1.1	We see a lot of images	iv
1.1.2	Cultural vs Individual Memorability	iv
1.1.3	How do we measure the memorability of an image	iv
1.1.4	Why do we want to create a memorable image	v
1.1.5	How does this research help to create a memorable image	v
1.2	Motivation	vi
1.3	What are my goals	vi
1.4	Structure	vi
2	Literature Review	vii
2.1	Memorability	vii
2.2	Using Deep Learning to Predict Memorability	ix
2.2.1	Memorability Datasets	ix
2.2.2	Image Synthesis Models	x
3	Methodology	xiii
3.1	Requirements Capture	xiii
3.2	Motivation	xiii
3.3	Experiment 1	xiv
3.4	Experiment 2	xiv
3.5	Experiment 3	xvi
3.6	Results Analysis/Testing	xvii
4	Experimental Results	xviii
4.1	Experimental Environment	xviii
4.2	Experiment 1	xviii
4.3	Experiment 2	xviii
4.4	Experiment 3	xxi
4.5	Comparisons	xxi
4.6	Results	xxi
4.7	Evaluation Metrics	xxi
4.8	Analysis	xxi
5	Conclusion	xxii

Contents

1 Introduction

1.1 Background

1.1.1 We see a lot of images

We are constantly surrounded by imagery, on the way to work, on the internet, on TV, in stores, etc. Some of them stick out more than others, we see hundreds, if not thousands, of images a day, and yet culturally and individually we all remember the same ones.

1.1.2 Cultural vs Individual Memorability

Due to cultural significance an image can become memorable. The 2015 dress [1], a country's flag, or the 1932 image "lunch atop a skyscraper" [2], come to mind. I would argue that the memorability of these images is tied to the culture surrounding them, not necessarily due to intrinsic properties within. The focus of this research is on the intrinsic properties within an image that make it memorable to an individual, like seeing an advert on a bus and then later recognising the same advert online.

1.1.3 How do we measure the memorability of an image

It has been shown that the memorability of an image is an intrinsic property, independent from the viewer [3]–[6]. This was achieved by performing a memorability game where participants are presented with a stream of images, and on an interval shown an image that they had already seen. Most participants were able, or unable, to remember the same images. How often an image was recognised is proportional to its memorability. This is taken further by Akagunduz et al. [7] where, in a similar experiment, they measure which regions of images are memorable.

1.1.4 Why do we want to create a memorable image

1.1.5 How does this research help to create a memorable image

1.2 Motivation

1.3 What are my goals

1.4 Structure

2 Literature Review

2.1 Memorability

In R. Breners work [An experimental investigation of memory span] subject's abilities to remember a series of units was tested, a unit was different depending on the test,

In the digit test, for example, each digit was a unit; in the sentence test each sentence was a unit, etc. ... Each nonsense syllable constituted a unit ... Each consonant constituted a unit ... Each [geometric] design constituted a unit.

It was found that people are not very good at remembering nonsense syllables or sentences but are much better at remembering digits, consonants, and colours. Remembering geometric designs is placed somewhere in the middle. This is of interest because it's very similar to our investigation into what aspects make images memorable. Brener was interested in consonants, colours, geometric designs, etc but the finding that different units can be significantly harder or easier to remember is useful to us. If we swap out units for image properties such as texture, contrast, saturation etc, or even more abstract features that a CNN may recognise, then it should be possible to find how memorability is impacted by these.

The mean score of 5.31 found for geometric designs seems to indicate that people are not great at recalling images, however in R. Nickerson's work [Short-term memory for complex meaningful visual configurations] we can see the opposite, subjects are found to be extremely good at recognising images. Nickerson found that subjects shown a series of images are, with great accuracy, able to recall if the photo is one they have seen before or not. An item is referred to as "new on its first occurrence and old on its second occurrence"[Short-term memory for complex meaningful visual configurations p.156], it was found that subjects shown one image at a time, each with equal chance of being old or new, are able to correctly distinguish them with 95% accuracy.

Like in [Short-term memory for complex meaningful visual configurations], R. Shephards work [Recognition Memory for Words, Sentences, and Pictures]

2 Literature Review

also found that subjects are able to distinguish with great accuracy new and old images, the mean percentage of images correctly identified was 99.7% after a delay 2 hours, 92.0% after a delay of 3 days. 87.0% after a delay of 7 days, and 57.7% after a delay of 120 days. The introduction of a delay into Nickerson's experiment allows us to see that regardless of whether the image is in our short-term memory (2 hour delay) or our long-term memory (3-120 days) subjects are able to correctly identify an incredible amount of images.

L. Standing built on the work of Nickerson, Shepard, and Brener in [Learning 10,000 pictures]. Four experiments were ran which tested memory capacity and retrieval speed for pictures and words, I am interested in the performance related to pictures specifically. He found that the capacity for image recognition from memory is almost limitless, when measured under testing conditions. In Standing's first experiment he tested 'normal' and 'vivid' images. He describes 'normal' images as those that "may be characterized as resembling a highly variegated collection of competent snapshots"[Learning 10,000 pictures, p.208], and 'vivid' images are described as "striking pictures ... with definitely interesting subject matter"[Learning 10,000 pictures, p.208]. Much like how in Brener's work [An experimental investigation of memory span] the memorability of an object was variable based on what classification of unit was tested, this work shows that on an even more specific level, the classification of the image into normal or vivid, has an impact on subjects memorability, Standing found that the vivid images were more likely to be remembered by subjects.

Lots of work has been done to further build on that of Standing, Nickerson, Brener et al. T. Brady et al. [Visual long-term memory has a massive storage capacity for object details] performed research into how much information is retained in long term memory and found that subjects are able to remember not just the gist of an image, but also fine grain information such as the state of objects within an image, and that they are able to distinguish between variants of objects shown in images. An example shown is an abacus in two different states, 13/14 subjects were able to distinguish which one they had seen before.

T. Konkle et al. build further on the work in [Learning 10,000 pictures] and [Visual long-term memory has a massive storage capacity for object details] by studying the impact that categorical distinctness has on memorability. This was tested by creating a dataset of images where composed of categories such as tables, cameras, and bread etc. Each category had between 1 and 16 images, a memory test like those performed in [Learning 10,000 pictures] and [Visual long-term memory has a massive storage capacity for object details] etc was performed and the percentage correctly identified was found to decrease as the number of images within a category increased. From this we can see that categorically distinct images are more

likely to be remembered.

Studies by Isola et al. [1, 11] were performed with the goal of identifying a collection of visual attributes that make images memorable, and to use those to predict the memorability of an image. It was found that properties such as mean saturation, and the log number of objects, has less impact on the memorability score than object statistics. Categories such as: person, person sitting, and floor, were most helpful image memorability. The categories: ceilings, buildings, and mountains, were least helpful. Their approach was limited by the fact that the object statistics were annotated by hand, this would both make automating the process of determining image memorability impossible, and limit them from finding any abstract properties that helped/hindered memorability.

Khosla et al. [4] built on the work in [1, 11] by, instead of determining memorability of an entire image, creating a model that discovers memorability maps of individual images without human annotation. These memorability maps are able to distinguish which areas of an image are remembered, forgotten, or hallucinated. Their approach, similar to Isola et al. in [1, 11], is limited by the arbitrarily picked list of features that define memorability.

2.2 Using Deep Learning to Predict Memorability

2.2.1 Memorability Datasets

Khosla et al. have created a dataset, LaMem [5], containing 60,000 annotated image and memorability pairs. They use deep learning to predict these memorability maps.

This is similar to the VISHEMA plus dataset [7], which I will be working with. In [7] the network output utilises fully connected layers which I believe is unnecessary, and may even hurt performance. More modern computer vision network structures, such as those in [8,9], use fully convolutional networks. These maintain spatial locality which allows them to generalise better.

In both [5,7] they present models that predict memorability maps of the images in the datasets presented, they do this by training a classifier to predict the memorability score of an image, and then to split the image into multiple sections which they then predict the memorability of producing a lower resolution memorability map.

I instead propose to use an Image-to-Image model such as those presented in [UNet paper, Pix2Pix paper, diffusion model paper] to take an image from the dataset and predict the memorability map of it from that.

Using an Image-to-Image model we should be able to predict memorability maps of images by training a fully convolutional network with a sufficiently large dataset such as those presented in [5,7].

2.2.2 Image Synthesis Models

I propose that predicting memorability maps is an extension of image segmentation, modern machine learning has found many network architectures that would be worth exploring for this problem

Recently, the use of deep convolutional networks [12,13] has proven exceptionally successful in image classification and object recognition. It stands to reason that a combination of these systems could perform well at this task. It could, in the worst case, recognise which objects are within an image, and to use that to help inform how memorable the image is. In the best case, it could learn to extract the features from categories that make them memorable and use this to predict the memorability of categories not yet seen. The weights within a convolutional layer act as filters to recognise patterns within images, through the stacking of these filters we are able to build networks that can recognise complex objects. The main limitation within [1,11,4] is that the features are arbitrarily picked by humans and manually annotated. The use of convolution should make it possible to learn these features automatically.

Auto-Encoders

The U-Net [8] has also found great success in image segmentation, this system however relies on the output being a classification. Something in between these two could work very well.

I will use a U-Net [U-Net paper] as my image to image network. The U-Net makes use of an Encoder block and a Decoder block, there are residual connections between different sections of the blocks that help to preserve image data in the decoding stage. The network has seen great success in the medical field. [U-net success paper].

Generative Adversarial Networks

Conditional GAN architectures have seen lots of success and produce high quality images, they work through the use of competitive co-evolutionary algorithms where a Generator and a Discriminator will compete. The Generator is typically given a latent space vector and uses that to produce an image, the Discriminator has to determine if that image belongs to a given distribution. In conditional GANS The weights of the two networks are updated through backpropagation based on if they win or lose. The Generator wins by fooling the Discriminator into predicting that the image is from the distribution. The Discriminator wins if it can accurately predict whether an image is from the distribution or not. These networks have seen great success but typically suffer from instability due to their adversarial nature, and often suffer from mode collapse. Despite this, lots of research has gone into producing GANs that produce high quality work.

Pix2Pix by Isola et al. [10], is an image-to-image deep learning architecture based on Conditional Generative Adversarial Networks. It is able to convert images from one style into another. For example, it can convert labels into a street scene or photos of daytime into night time. It makes use of a UNet backbone and has found great success.

Diffusion Models

Diffusion based image generation is inspired by nonequilibrium thermodynamics, it works through a forward diffusion process of adding Gaussian noise to a sample until it has been transformed such that it is no longer distinguishable to Gaussian noise, this should follow a diffusion schedule where noise is applied T times. We then train a model to compute a reverse diffusion process. That is, given a value $1 < t < T$, and a sample that has been transformed t times, predict the sample at $t-1$ transformations. When we want to produce a sample, we create $x_0 \sim N(0, 1)$ and apply reverse diffusion to it T times, following our diffusion schedule.

Diffusion models were first introduced by J. Sohl-Dickstein et al. (Deep Unsupervised Learning using Nonequilibrium Thermodynamics) as a flexible and tractable machine learning model. They describe the forward diffusion process of systematically destroying the structure in a data distribution and the learning of a backwards process to restore the structure. The method uses a Markov chain to convert x_t into x_{t-1} . Starting with x_T , a sample of Gaussian noise, a generative Markov chain converts this into x_0 , which is a sample from the target data distribution. Because the model only estimates small perturbations of noise, x_t given x_{t-1} , rather than an entire transformation from x_0 to x_T , it is tractable to train.

2 Literature Review

This was then built on by Y. Song and S. Ermon in (Generative Modeling by Estimating Gradients of the Data Distribution) by using Langevin dynamics

The DDPM a diffusion model based on a UNet, introduced by J. Ho et al. (Denoising Diffusion Probabilistic Models) is capable of producing high quality images, they show that it is capable of achieving state of the art FID scores across the CIFAR10 dataset. Dhariwal and Nichol (Diffusion Models Beat GANs on Image Synthesis) show tweaks that allow for diffusion models to achieve state of the art FID scores across the ImageNet dataset and when used in combination with upsampling diffusion further improve FID scores. They do this by using improvements proposed by Song et al. [Denoising diffusion implicit models], Nichol and Dhariwal [Improved denoising diffusion probabilistic models], Song et al. [Score-based generative modeling through stochastic differential equations], [Large scale gan training for high fidelity natural image synthesis] [A style based generator architecture for generative adversarial networks]. Through these improvements they are able to improve image quality while reducing the number of noise steps from 1000 to (in some cases) 50. Through the decrease in noise steps they are able to reduce the amount of time that it takes to generate an image.

Chen discusses in [On the importance of noise scheduling] how changing the resolution of an image has an impact on the noise scheduling required, he finds that at a smaller resolution the optimal schedule may cause under training for higher resolution images. Multiple strategies are proposed to adjust noise scheduling. Firstly, changing the noise schedule functions to those based on cosine or sigmoid, with temperature scaling. Secondly, reducing the input scaling factor from 1 increases the noise levels which destroys more information at the same noise level. They then combine these into a compound noise scheduling strategy.

3 Methodology

3.1 Requirements Capture

The VISHEMA dataset[Vischema paper] contains image to vms mappings. I aim to use a deep learning model to learn a mapping of these images to their corresponding labels. Our model should learn a general understanding of the mapping such that when it is provided with an image that matches our distribution, it can accurately create a VMS label for it. Our model will need to learn to create accurate mappings and we can test that through a loss function, such as L1, and through qualitative analysis.

3.2 Motivation

As standard backpropagation, GANs, and diffusion all produce images in different ways I think it would be interesting to compare how the 3 of them perform when asked the same task. I will run three experiments: I will train a UNet, a GAN, and a diffusion model to produce a VMS label given an image from the dataset. I hope to be able to find the strengths and weaknesses of each of these systems in this application.

For each system I will experiment with network parameters and training hyperparameters to fine tune models that produce the best output. I will experiment with training the models to produce the sum of the images and the VMS labels, and with training the models to produce just the VMS label. In the former the VMS label can be calculated from the output.

In each experiment I will automate a system to: test the use of different values for the number of layers and channels in each network, vary the normalisation method, optimisation function, learning rate, and batch size to find the best training environment and model parameters. As it is not be feasible to test every combination of these variables, I will employ a training strategy to test every variation of a single parameter/hyperparameter while keeping the rest constant, pick the best scoring variation of that parameter, and move on to the next. I will start with choices that I estimate to be good,

as this should only improve on them. This will bring the size our search space down by an order of magnitude, however we won't be exploring the entire search space and will potentially miss out on good values. This will be especially necessary in experiment 2 as training within a GAN is typically unstable [papers showing instability in GANs]. In experiment 3 I will also vary the noise scheduler and the beta values.

3.3 Experiment 1

Model: This network is a UNet autoencoder that takes as input a 64x64x3 tensor of floating point values in the range [-1,1], these are the images in our dataset. It outputs a 64x64x3 tensor of floating point values in the range [-1,1], these are our label estimates. This model can be described as:

$$L_{pred} = model(I)$$

Model Loss: This is simply the L1 loss between L_{pred} and L_{real} , describing how closely the output of our network matches the corresponding label. At each epoch we will calculate this over the training dataset, use that for backpropagation, and then calculate it across the validation dataset to test how well we have generalised.

$$L = L1(model(I), L_{real})$$

Training Strategy: Please see algorithm 1 for the training loop.

Algorithm 1 UNet Autoencoder Training Strategy

```

1: for every epoch do
2:   for  $I, L_{real}$  in training dataloader do
3:
4:      $L_{pred} = model(I)$ 
5:      $loss = L1(L_{pred}, L_{real})$ 
6:     Update weights of model with backpropagation
7:
8:   end for
9: end for

```

3.4 Experiment 2

I will train a conditional generative adversarial network to generate images of VMS maps given images from the VISHEMA dataset. The Generator

3 Methodology

will implemented using a UNet and the Discriminator will be a fully convolutional PatchGAN classifier as described by Isola et al. in [Pix2Pix paper], producing a 4x4 array of boolean values, describing whether a region of the input image is from the dataset, or if it is generated.

Generator: This network takes as input a 64x64x3 tensor of floating point values in the range [-1,1] It outputs a 64x64x3 tensor of floating point values in the range [-1,1]. This model can be described as:

$$L_f = G(I)$$

Discriminator: This network takes as input a 64x64x6 tensor of floating point values in the range [-1,1]. The first 3 channels are an image and the second 3 channels are its corresponding label, either real or fake. It outputs a 4x4x1 tensor of boolean values. Each value in this output represents a 16x16x3 region of the input. This model can be described as:

$$\{0|1\}^{16} = D(L)$$

The loss for each network is computed as described in [pix2pix paper]. The loss is computed across an entire batch of images and then the weights are adjusted with backpropagation.

Generator Loss: The generator loss describes how well it can trick the discriminator, and how closely its output matches the real label for the given image.

$$L = \text{MSE}(D(L_f, I), 1) + \text{L1}(L_f, L_r)$$

Discriminator Loss: When training the Discriminator the loss is calculated as the mean value of how close our prediction was to the correct answer across a fake and real label.

$$L = 0.5 \times (\text{MSE}(D(L_f, I), 0) + \text{MSE}(D(L_r, I), 1))$$

Because these two loss values are relative to the performance of eachother they can't be used to see if our generator has converged on a solution. Therefore it is necessary, at each epoch, to also compute the L1 loss of the fake labels and real labels across the validation dataset, as this value is independent of the discriminator. This will inform me of whether the generator is overfitting, underfitting, or training well.

Training Strategy: Please see algorithm 2 for the training loop.

Algorithm 2 GAN Training Strategy

```

1: for every epoch do
2:   for  $I, L_r$  in training dataloader do
3:
4:      $L_f = G(I)$ 
5:
6:      $pred_{fake} = D(L_f, I)$ 
7:      $loss_G = MSE(pred_{fake}, 1) + L1(L_f, L_r)$ 
8:     Update weights of G with backpropagation
9:
10:     $pred_{real} = D(L_r, I)$ 
11:     $loss_D = 0.5 \times (MSE(pred_{fake}, 0) + MSE(pred_{real}, 1))$ 
12:    Update weights of D with backpropagation
13:
14:   end for
15: end for

```

3.5 Experiment 3

Model: I will train a diffusion-based model to generate images of VMS maps given images from the VISHEMA dataset. It will be implemented using a UNet. This network takes as input a 64x64x6 tensor of floating point values, this is the concatenation of an image and the noise added after t steps. It outputs a 64x64x3 tensor of floating point values, this is the models estimate of the noise added at time $t - 1$. This model can be described as

$$noise_{pred} = model(I, t)$$

Model Loss: The model loss describes how well it can estimate the noise added to an image between time steps $t - 1$ and t , it is computed across an entire batch of images and then the model weights are adjusted with backpropagation.

$$L = MSE(model(I, t), noise_{real})$$

The loss calculated for backpropagation is relative to the performance in estimating noise added, not for the performance when calculating VMS labels. Because of this I will also have to calculate the L1 loss of the generated labels against the real labels. This will take a lot of computational time so I will do it at the end of training. I will be able to see if our model has converged by observing the backpropagation loss.

Training Strategy: Please see algorithm 3 for the training loop.

Algorithm 3 Diffusion Model Training Strategy

```

1: for every epoch do
2:   for  $I, L$  in training dataloader do
3:
4:      $t = \text{random}(0, \text{noise\_steps})$ 
5:      $\text{noise}_{\text{real}}, L_{\text{noisy}} = \text{noise\_image}(L, t)$     ▷ Apply  $t$  steps of noise
        and return noise added at step  $t$ 
6:
7:      $\text{input} = \text{concatenation}(I, L_{\text{noisy}})$ 
8:      $\text{noise}_{\text{pred}} = \text{model}(I, t)$ 
9:      $\text{loss} = \text{MSE}(\text{noise}_{\text{pred}}, \text{Noise}_{\text{real}})$ 
10:    Update weights of model with backpropagation
11:  end for
12: end for
  
```

3.6 Results Analysis/Testing

After performing all 3 experiments I will compare the results across the validation dataset qualitatively and using the L1 loss. The L1 loss will tell me how statistically close the outputs are but through qualitative analysis I can see if the outputs have cheated. I can also compare the L1 loss across the training and validation sets to see if any models have generalised well, if the loss across the training dataset is much smaller/larger than the validation dataset then it will imply that the model has become overfit/underfit respectively. Ideally they should be similar.

4 Experimental Results

4.1 Experimental Environment

I have implemented the experiments in Python using the PyTorch machine learning framework, however the methodology that I describe should produce the same results in any programming language or framework.

This model was trained on a computer using an RTX 3070 with 8GB of VRAM and an AMD Ryzen 3600 with 32GB of system RAM. Testing the hyperparameter options for experiment 1 took approximately X days and I trained the final model over the course of Y hours. Testing the hyperparameter options for experiment 2 took approximately 3 days and I trained my final model over the course of 5 hours. Testing the hyperparameter options for experiment 1 took approximately X days and I trained the final model over the course of Y hours.

4.2 Experiment 1

4.3 Experiment 2

In order to find which parameter and hyperparameter combinations would perform the best I automated a training process which would vary the following:

1. The normalisation layer for both the generator and the discriminator.
2. The channel layouts for the generator and the discriminator.
3. The optimiser used for the generator and the discriminator.
4. The learning rates used for each optimiser.
5. If the Adam optimiser was found to be ideal for either the generator or discriminator, then to use different beta options.

I explored the following options:

Normalisation layers:

- Batch Normalisation
- Instance Norm

Generator channel layouts:

- encoder:(3, 64, 128, 256, 512, 1024), decoder:(1024, 512, 256, 128, 64),
- encoder:(3, 32, 64, 128, 256, 512), decoder:(512, 256, 128, 64, 32),
- encoder:(3, 50, 100, 200, 400, 800), decoder:(800, 400, 200, 100, 50),
- encoder:(3, 100, 200, 400, 800, 1600), decoder:(1600, 800, 400, 200, 100)

Discriminator channel layouts:

- (6, 64, 128, 256, 512, 1024),
- (6, 32, 64, 128, 256, 512),
- (6, 50, 100, 200, 400, 800),
- (6, 100, 200, 400, 800, 1600)

Optimisers:

- SGD, using the following learning rates (0.005, 0.01, 0.02)
- Adam, using the following learning rates (0.0005, 0.001, 0.002), and using the following betas ((0.9, 0.999), (0, 0.999), (0.5, 0.999))
- Adadelta, using the following learning rates (0.5, 1, 2)

Exploring this search space exhaustively is unfortunately not feasible, there are over 5000 different combinations possible, and if I tested each combination once for 100 epochs then it would take approximately 300 days to test. Instead I will have to explore a subset of this search space. I estimated some good default parameter and hyperparameter values, by varying these values I can lower the scope of the search space to approximately 100 combinations, which took ~ 3 days to test. Unfortunately this did mean that every combination hasn't been tested, however it should give us a good

4 Experimental Results

approximation of the best combination.

Note: I didn't need to specify default values for the normalisation layer as these were the first variables I tested.

Default Generator Parameters and Hyperparameters:

- Encoder Channel Layout: (3,64,128,256,512,1024)
- Decoder Channel Layout: (1024,512,256,128,64)
- Optimiser: Adam, betas = (0.9, 0.999)
- Learning Rate: 0.001

Default Discriminator Parameters and Hyperparameters:

- Channel Layout: (6,64,128,256,512,1024)
- Optimiser: Adam, betas = (0.9, 0.999)
- Learning Rate: 0.001

I then iterated over each different parameter and hyperparameter and varied each one sequentially. I tested each combination for 100 epochs and used the L1 loss across the validation dataset as the score, if any parameter options performed better than the default then they would be used going forward. This meant that I only had to iterate ~ 100 combinations. This took 3 days to test, and at the end I found that the following combination was the best:

Best Generator Parameters and Hyperparameters:

- Normalisation Layer: Batch Normalisation
- Encoder Channel Layout: (3,32,64,128,256,512)
- Decoder Channel Layout: (512, 256, 128, 64,32)
- Optimiser: SGD
- Learning Rate: 0.01

Best Discriminator Parameters and Hyperparameters:

- Normalisation Layer: Batch Normalisation
- Channel Layout: (6,100,200,400,800,1600)

4 *Experimental Results*

- Optimiser: Adam, betas = (0.9, 0.999)
- Learning Rate: 0.001

With this combination we achieved a loss of ~ 1.02 across the validation dataset. I found other good results using similar combinations. Using the Adam optimiser for both the generator and discriminator achieved a loss of ~ 1.04 across the validation dataset. Using the Adatelta optimiser for the generator and the Adam optimiser for the discriminator achieved a loss of ~ 1.03 across the validation dataset.

4.4 Experiment 3

4.5 Comparisons

Varying Normalisation

Varying Optimiser

Varying Loss Function

PatchGAN vs Boolean Discriminator

Benefits of Residual Connections

4.6 Results

4.7 Evaluation Metrics

4.8 Analysis

5 Conclusion

Bibliography

- [1] BBC. 'Optical illusion: Dress colour debate goes global.' (2015), [Online]. Available: <https://www.bbc.com/news/uk-scotland-highlands-islands-31656935>.
- [2] M. Gambino. 'Lunch atop a skyscraper photograph: The story behind the famous shot.' (2012), [Online]. Available: <https://www.smithsonianmag.com/history/lunch-atop-a-skyscraper-photograph-the-story-behind-the-famous-shot-43931148/>.
- [3] P. Isola, J. Xiao, A. Torralba and A. Oliva, 'What makes an image memorable?' In *IEEE Conference on Computer Vision and Pattern Recognition (CVPR)*, 2011, pp. 145–152.
- [4] P. Isola, D. Parikh, A. Torralba and A. Oliva, 'Understanding the intrinsic memorability of images,' in *Advances in Neural Information Processing Systems*, 2011.
- [5] A. Khosla, A. S. Raju, A. Torralba and A. Oliva, 'Understanding and predicting image memorability at a large scale,' in *International Conference on Computer Vision (ICCV)*, 2015.
- [6] P. Isola, J. Xiao, D. Parikh, A. Torralba and A. Oliva, 'What makes a photograph memorable?' *Pattern Analysis and Machine Intelligence, IEEE Transactions on*, vol. 36, no. 7, pp. 1469–1482, 2014.
- [7] E. Akagunduz, A. G. Bors and K. K. Evans, 'Defining image memorability using the visual memory schema,' *IEEE Transactions on Pattern Analysis and Machine Intelligence*, vol. 42, no. 9, pp. 2165–2178, 2020. DOI: 10.1109/TPAMI.2019.2914392.